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MATHÉMATIQUES

SUR LES SOLUTIONS PÉRIODIQUES DES ÉQUATIONS DIFFÉRENTIELLES NON-LINÉAIRES DU TYPE PARABOLIQUE

Par Dj. Kh. KARIMOV

(Présenté par S. L. Soboleff, de l'Académie, le 15. VII. 1939)

Dans le présent travail nous considérons l'équation différentielle non-linéaire

$$\frac{\partial z}{\partial t} - a^2 \frac{\partial^2 z}{\partial x^2} = \Phi(x, t) + \mu f(z). \tag{1}$$

Il s'agit de trouver la solution de l'équation (1) satisfaisant aux conditions

$$z(0, t) = z(1, t) = 0; \quad z(x, 0) = z(x, 1). \tag{2}$$

Ce problème aux limites sera résolu par une méthode d'approximations successives. Supposons que la fonction $\Phi(x, t)$ satisfait à une condition de Lipschitz

$$|\Phi(x_1, t_1) - \Phi(x_2, t_2)| \leqslant M_1 |x_1 - x_2| + M_2 |t_1 - t_2|$$

pour tous les points (x_1, t_1), (x_2, t_2) appartenant au domaine

$$\overline{D} = D \begin{pmatrix} 0 \leqslant x \leqslant 1 \\ 0 \leqslant t \leqslant 1 \end{pmatrix}$$

et que sa dérivée partielle $\frac{\partial \Phi}{\partial x} = \Phi_1(x, t)$ est une fonction continue et à variation bornée relativement à la variable x, pour tout $0 \leqslant t \leqslant 1$, et relativement à la variable t, pour tout $0 \leqslant x \leqslant 1$. Nous supposons encore que $\Phi(x, t)$ est une fonction périodique de la variable t, de période égale à un et développable en série de Fourier de sinus:

$$\Phi(x, t) = \sum_{n=1}^{\infty} \Phi_n(t) \sin n\pi x,$$

où

$$\Phi_n(t) = 2 \int_0^1 \Phi(\eta, t) \sin n\pi\eta \, d\eta.$$

Nous appellerons conditions (A) les conditions auxquelles nous venons d'assujettir la fonction $\Phi(x, t)$.

Théorème. *L'équation différentielle* (1) *admet une solution unique continue, avec ses dérivées partielles du second ordre* *, *dans le domaine* $\overline{D}$, *satisfaisant aux conditions* (2) *et périodique relativement à* t, *de période égale à* 1, *à condition que:*

1) *La fonction* $\Phi(x, t)$ *satisfasse aux conditions* (A).

2) $|f'(z_1) - f'(z_2)| \leqslant M |z_1 - z_2|$.

3) $f(z) = -f(-z)$; $f(0) = 0$.

4) $|\mu| < \frac{3a^2}{2N}$,

où $N = \sup |f'(z)|$ *dans un certain domaine fini de la variable* z, *et* a *étant le paramètre de l'équation* (1).

Nous allons résoudre ce problème par des approximations successives. Pour l'approximation initiale (d'ordre zéro) $z_0(x, t)$ nous prenons la solution de l'équation

$$\frac{\partial z_0}{\partial t} - a^2 \frac{\partial^2 z_0}{\partial x^2} = \Phi(x, t) \tag{1'}$$

satisfaisant aux conditions (2). La solution de cette équation sera de la forme:

$$z_0(x, t) = \sum_{n=1}^{\infty} T_n^0(t) \sin n\pi x \tag{3}$$

où

$$T_n^0(t) = 2 \int_0^t d\xi \int_0^1 \Phi(\eta, \xi)\, e^{-a^2 n^2 \pi^2 (t-\xi)} \sin n\pi\eta d\eta + \\ + \frac{2}{1 - e^{-a^2 n^2 \pi^2}} \int_0^1 d\xi \int_0^1 \Phi(\eta, \xi)\, e^{-a^2 n^2 \pi^2 (t-\xi+1)} \sin n\pi\eta d\eta.$$

On voit facilement que $z_0(x, t)$ est une fonction continue avec ses dérivées partielles du second ordre dans le domaine $\overline{D}$, périodique en t et de période égale à 1. L'approximation d'ordre $k-1$, $z_{k-1}(x, t)$ étant donnée, nous définissons l'approximation d'ordre k, $z_k(x, t)$ comme solution de l'équation

$$\frac{\partial z_k}{\partial t} - a^2 \frac{\partial^2 z_k}{\partial x^2} = \Phi(x, t) + \mu f(z_{k-1})$$

satisfaisant aux conditions (2). Cette approximation sera de la forme

$$z_k(x, t) = \sum_{n=1}^{\infty} T_n^k(t) \sin n\pi x, \tag{4}$$

où

$$T_n^k(t) = 2 \int_0^t d\xi \int_0^1 [\Phi(\eta, \xi) + \mu f(z_{k-1})]\, e^{-a^2 n^2 \pi^2 (t-\xi)} \sin n\pi\eta d\eta + \\ + \frac{2}{1 - e^{-a^2 n^2 \pi^2}} \int_0^1 d\xi \int_0^1 [\Phi(\eta, \xi) + \mu f(z_{k-1})]\, e^{-a^2 n^2 \pi^2 (t-\xi+1)} \sin n\pi\eta d\eta.$$

La fonction $z_k(x, t)$ est une fonction continue avec ses dérivées $\frac{\partial^2 z_k}{\partial x^2}$, $\frac{\partial z_k}{\partial t}$ dans le domaine $\overline{D}$. La limite

$$\lim_{k \to \infty} z_k(x, t) = z(x, t)$$

* Savoir: la dérivée première en t et la dérivée seconde en x.

existe et cette limite fournit la solution du problème posé. Pour le démontrer il suffit de considérer la série

$$z = z_0 + (z_1 - z_0) + \ldots + (z_k - z_{k-1}) + \ldots; \tag{5}$$

cette série converge absolument et uniformément dans le domaine $\overline{D}$. En effet

$$|z_k - z_{k-1}| \leqslant 2|\mu| \sum_{n=1}^{\infty} \left[\int_0^t e^{-a^2n^2\pi^2(t-\xi)}\, d\xi \int_0^1 [f(z_{k-1}) - f(z_{k-2})]\, d\eta + \right.$$

$$\left. + \frac{1}{1 - e^{-a^2n^2\pi^2}} \int_0^1 e^{-a^2n^2\pi^2(t-\xi+1)}\, d\xi \int_0^1 [f(z_{k-1}) - f(z_{k-2})]\, d\eta \right] \leqslant$$

$$\leqslant 4|\mu| N \sup |z_{k-1} - z_{k-2}| \sum_{n=1}^{\infty} \frac{1}{a^2n^2\pi^2} = \frac{2|\mu| N \sup |z_{k-1} - z_{k-2}|}{3a^2}$$

d'où

$$\sup |z_k - z_{k-1}| \leqslant \frac{2|\mu| N}{3a^2} \sup |z_{k-1} - z_{k-2}|. \tag{6}$$

En posant dans la formule (6) $k = 2$, $k = 3, \ldots$, nous aurons:

$$|z_2 - z_1| \leqslant \frac{2|\mu| N}{3a^2} \sup |z_1 - z_0|$$

. . . . . .

$$|z_k - z_{k-1}| \leqslant \left[\frac{2|\mu| N}{3a^2} \right]^{k-1} \sup |z_1 - z_0|.$$

. . . . . .

Donc, la série convergente à termes constants

$$|z_0| + L + L\left[\frac{2|\mu| N}{3a^2}\right] + \ldots + L\left[\frac{2|\mu| N}{3a^2}\right]^n + \ldots, \quad [L \geqslant \sup |z_1 - z_2|]$$

est une série majorante de la série (5). Par conséquent, la série (5) converge absolument et uniformément dans le domaine $\overline{D}$.

Par conséquent, nous pouvons écrire

$$z(x,t) = 2\sum_{n=1}^{\infty} \left[\sin n\pi x \int_0^t e^{-a^2n^2\pi^2(t-\xi)}\, d\xi \int_0^1 [\Phi(\eta,\xi) + \mu f(z)] \sin n\pi\eta\, d\eta + \right.$$

$$\left. + \frac{1}{1 - e^{-a^2n^2\pi^2}} \int_0^1 e^{-a^2n^2\pi^2(t-\xi+1)}\, d\xi \int_0^1 [\Phi(\eta,\xi) + \mu f(z)] \sin n\pi\eta\, d\eta \right].$$

On voit facilement que les séries

$$\frac{\partial^2 z}{\partial x^2} = \sum_{k=0}^{\infty} \left[\frac{\partial^2 z_k}{\partial x^2} - \frac{\partial^2 z_{k-1}}{\partial x^2} \right]; \quad \frac{\partial z}{\partial t} = \sum_{k=0}^{\infty} \left[\frac{\partial z_k}{\partial t} - \frac{\partial z_{k-1}}{\partial t} \right] \quad [z_{-1} = 0]$$

sont, elles aussi, absolument et uniformément convergentes dans le domaine $\overline{D}$. Il reste à démontrer l'unicité de la solution trouvée $z(x,t)$.

Supposons qu'il existe deux solutions continues distinctes z et v de l'équation (1), satisfaisant aux conditions (2); alors la différence $z - v = k$ sera une solution de l'équation différentielle

$$\frac{\partial k}{\partial t} - a^2 \frac{\partial^2 k}{\partial x^2} = \mu [f(z) - f(v)]$$

aussi bien qu'une solution de l'équation intégrale

$$k(x,t) = 2\mu \sum_{n=1}^{\infty} \sin n\pi x \int_0^t e^{-a^2 n^2 \pi^2 (t-\xi)}\, d\xi \int_0^1 [f(z) - f(v)] \sin n\pi\eta\, d\eta +$$

$$+ 2\mu \sum_{n=1}^{\infty} \frac{\sin n\pi x}{1 - e^{-a^2 n^2 \pi^2}} \int_0^1 e^{-a^2 n^2 \pi^2 (t-\xi+1)}\, d\xi \int_0^1 [f(z) - f(v)] \sin n\pi\eta\, d\eta.$$

En évaluant la valeur absolue de k, nous aurons

$$\sup |z - v| \leqslant \frac{2|\mu| N \sup |z - v|}{3a^2}$$

c'est-à-dire

$$|\mu| \geqslant \frac{3a^2}{2N}$$

ce qui est impossible, donc

$$z \equiv v$$

et le théorème est démontré.

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PHYSICS

SCATTERING OF PHOTONEUTRONS FROM DEUTERIUM BY NUCLEI OF ATOMS OF HEAVY ELEMENTS

By T. GOLOBORODKO and A. LEIPUNSKI

(Communicated by S. I. Wawilow, Member of the Academy, 3. VIII. 1939)

In this work we proceed with our measurements of the scattering cross-sections of RaTh-D photoneutrons with energies of about 210 kV with nuclei of atoms of light elements, from hydrogen to calcium.

The procedure was as previously described. 125 mg of RaTh in a small glass ampula were placed in the center of a thin-walled brass sphere filled with 50 cm³ of 99.6% D_2O. A cylindrical dysprosium detector surrounded with a paraffin sphere 13 cm in diameter was placed 27 cm from the ampula. The detector was irradiated to saturation whereupon its activity was measured with a thin-walled Geiger-Müller counter. The scatterer in the form of a cylindrical disc 8 cm in diameter was placed 9 cm from the center of the source. The measurements were made immediately one after another with and without the scatterer. The background of the counter was not checked during the measurements of activity of the detector; it was determined separately after the measurement of the detector was finished. The initial activity of the detector was 90 kicks per minute, the background being 10 kicks per minute*.

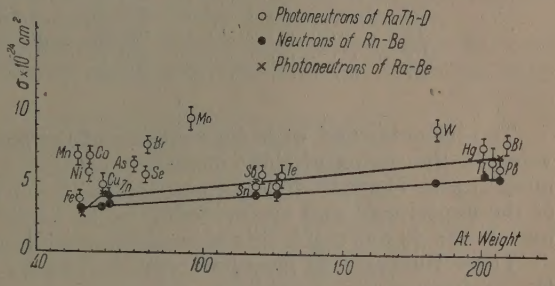

Fig. 1.

The sensitivity of the counter was checked by a uranium standard.

Correction was made for the non-paralled beam of photoneutrons (the geometry of the arrangement). The cross-sections σ were computed by the formula

$$\sigma = \frac{1}{N\delta} \ln \frac{1-\alpha}{\frac{a}{b}-\alpha} \qquad (1)$$

* In order to reduce the background, the counter was shielded with a layer of lead 10 cm thick.

where α is the correction for the geometry of the arrangement; N, the number of nuclei in a cubic centimeter of the scatterer; δ, the thickness of the scatterer; a and b, the numbers of kicks with and without the scatterer, respectively.

The results of the measurements are given in the table and the figure. The scattering cross-sections of Ra-Be photoneutrons previously measured in this laboratory are given for comparison. Moreover, in the figure are shown Dunning's[1] data for fast Rn-Be neutrons.

Element	Atomic weight	$\sigma \times 1$ ^{24}cm^2 Photoneutrons RaTh + D	$\sigma \times 10^{24}$cm^2 Photoneutrons Ra + Be
Mn	54.93	6.9±0.7	
Fe	55.84	3.6±0.3	2.7±0.5
Co	58.94	6.8±0.6	
Ni	58.69	5.7±0.6	
Cu	63.57	4.8±1.0	4.2±0.7
Zn	65.38	4.0±0.4	4.0±0.7
As	74.93	6.3±0.5	
Se	79.20	5.5±0.5	
Br	79.92	7.8±0.8	
Mo	96.00	9.6±1.2	
Sn	118.70	4.9±0.6	5.2±1.1
Sb	121.76	5.8±0.7	
Te	127.50	5.7±0.9	
I	126.93	4.7±0.8	
W	184.00	9.0±0.7	
Hg	200.61	7.8±0.6	
Tl	204.39	6.7±1.2	
Pb	207.22	6.2±0.7	7.3±1.4
Bi	209.00	8.4±0.8	

The irregular and wide fluctuations of the scattering cross-sections discovered in the region of light elements continue in the region of heavier elements up to bismuth. These fluctuations lie far beyond the statistical error of the experiment and are evidently to be ascribed to the properties of the nuclei. We do not think this phenomenon can be easily explained at present.

These fluctuations may be accounted for by our measuring not exactly the scattering cross-section, but a summary cross-section of scattering + absorption. However, this explanation must probably be discarded, since the absorption cross-sections of many elements with photoneutrons of the given energies were measured by Griffiths[2] and by Halban and Kowarski[3], while the upper limits of the absorption cross-sections were determined in this laboratory[4]. All these measurements yield small absorption cross-sections of the order of 10^{-25}—10^{-26} cm^2.

For closer investigation of these phenomena we are now measuring the scattering cross-sections of photoneutrons of other energies.

Ukrainian Physical-Technical Institute.
Kharkov.

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1. VIII. 1939.

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PHYSIQUE

OBSERVATIONS SUR LA DÉSINTÉGRATION DES NOYAUX DE L'URANE SITUÉS À L'INTÉRIEUR DE L'ÉMULSION DES PLAQUES PHOTOGRAPHIQUES À COUCHE ÉPAISSE ET BOMBARDÉS PAR DES NEUTRONS

Par A. P. JDANOFF et L. W. MYSSOWSKY

(Présenté par V. G. Chlopine, de l'Académie, le 15. VIII. 1939)

Nous avons déjà communiqué dans les Notes précédentes([1],[2]) les résultats des expériences sur l'observation de la désintégration de l'urane sous l'action des neutrons. On a employé dans les expériences suivantes les plaques à l'émulsion spéciale, dans laquelle on obtenait les traces des noyaux de recul essentiellement plus épaisses que celles des particules α.

Mais, comme il y avait au-dessus de la plaque une couche épaisse de l'urane, les parcours de la majorité des particules étaient petits. Pour cela, dans

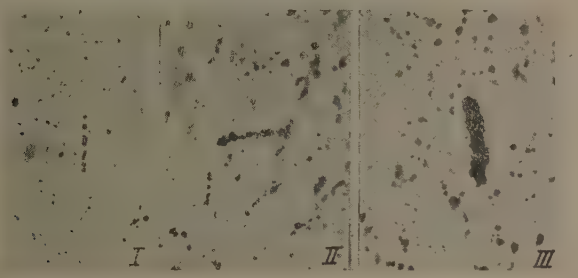

Fig. 1. Fig. 2. Fig. 3.

certains cas, il était difficile de reconnaître les traces des noyaux de recul parmi celles des particules α. Afin d'éviter ces difficultés, nous avons effectué des expériences avec des plaques, dans la couche d'émulsion desquelles on a directement introduit des sels d'urane. On l'obtenait en baignant les plaques photographiques à couche épaisse (50 μ) et à l'émulsion spéciale pendant environ 20 minutes dans une solution d'uranile, de teneur 10^{-4} g/cm^3. Ensuite on faisait sécher les plaques dans le courant de l'air chaud, après quoi on les mettait immédiatement sous les rayons de neutrons obtenus par un cyclotron au moyen de la réaction ${}^2_1D + {}^2_1D \rightarrow {}^3_2He + {}^1_0n$. On développait les plaques et on les étudiait sous un microscope à l'oculaire stéréoscopique. On développait en même temps les plaques de contrôle avec le sel d'urane introduit, mais non bombardées par des neutrons.

L'examen a montré que les plaques de contrôle n'ont que quelques traces des particules α de l'urane dont les parcours dans l'air sont en majeure partie de 2.5 cm. Sur les plaques irradiées par des neutrons on observait non seulement des traces des particules α de l'urane, mais encore des traces plus épaisses et un peu plus longues, que nous croyons devoir attribuer à la désintégration des noyaux de l'urane sous l'action de bombardement par des neutrons. Les microphotographies montrent le caractère des traces observées. La fig. 1 donne la trace d'une particule α de l'urane (ayant un parcours dans l'air de 2.5 cm) sur une plaque de contrôle. La fig. 2 montre le type le plus fréquent de la désintégration des noyaux de l'urane, le parcours total des noyaux de recul partant dans les sens opposés étant égal, dans ce cas, à 3.6 cm dans l'air. Enfin la fig. 3 présente la désintégration de noyau de l'urane aux traces très épaisses, le parcours total est 4.0 cm dans l'air.

Il est possible que ce cas correspond à un procédé de «désintégration», tel qu'on obtient des noyaux de recul à la charge effective supérieure. Le rapport du nombre des cas semblables relativement au nombre des cas représentés sur la fig. 2 est environ 1 : 1000. La grandeur de la section effective d'après les données de nos expériences est voisine de 10^{-24} cm^2.

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Leningrad.

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PHYSICS

ON THE MECHANISM OF SECONDARY EMISSION OF ELECTRONS FROM COMPOSITE SURFACES

By P. V. TIMOFEYEV

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The theory of secondary emission from metals was given in the works of Fröhlich(1) and Nadyshevich. The latter explained very clearly all the processes which take place during the bombardment of the surface of a metal by a beam of electrons.

However, at the present time even definite qualitative conceptions do not exist concerning the processes taking place during secondary emission from composite surfaces of the type of an oxygen-caesium emitter. A number of authors, Morgulis(2), Khlebnikov(3), Bruining(4) and others explain the large values of the coefficient of secondary emission from these emitters as being mainly due to the relatively low concentration of conductivity electrons in the intermediate layer of such emitters. In the crystalline lattice of a semi-conductor where the concentration of conductivity electrons is small, the length of the mean free path of the electrons should, in the opinion of these authors, be considerably greater than the length of the free path of secondary electrons in a metal. The reduction in the length of the free path of secondary electrons with increase in the concentration of conductivity electrons is due to Coulomb interaction between the secondary and conductivity electrons. Semi-conductors contain bound electrons with low bond energy instead of conductivity electrons, in approximately the same concentrations as these latter are found in metals.

The energy loss of a secondary electron on movement in the intermediate layer of an oxygen-caesium emitter, owing to interaction between it and electrons with low bond energy, was evaluated by Morgulis on the basis of conceptions of impact ionization; this evaluation is clearly incorrect.

On interaction between a conductivity electron and a bound one the energy loss is:

$$\Delta W_2 = \frac{Q + E_0'}{\lambda}$$

where Q is the bond energy; E_0', the initial energy of a tertiary electron, which arises as a result of interaction between the secondary and the bound electron; λ is the length of the ionizing collision. For the case of an oxygen-caesium cathode Morgulis supposes that $Q = 2—3$ eV. which is completely untrue, since values of Q less than 1 eV are obtained by elementary calculation from comparison of the spectral characteristics and the energy distribution curves of photoelectrons(5).

It will easily be seen that under these conditions the dispersion of the energy of a secondary electron within the intermediate layer of the oxygen-caesium emitter will differ very little from the energy dispersion of a secondary electron during its movement in a metal.

In general it is highly probable that in semi-conductors the energy dispersion of the secondary electrons by weakly bound electrons will be of the same order of magnitude as the energy dispersion of secondary electrons caused by conductivity electrons in a metal, and consequently, the length of the free path of the secondary electrons should be of the same order in both cases.

Also, in semi-conductors the magnitude of the secondary emission may be greater than the secondary emission from metals as a consequence of the circumstance that in semi-conductors, in distinction from what happens in metals, a normal incidence of the primary electron beam can cause part of the secondary electrons at the moment of their formation to obtain an increment of impulse in the direction of the surface.

The magnitude and direction of the velocity of a secondary electron depend on the conditions, in which the electron was at the instant of receiving energy from the primary electron.

In semi-conductors and insulators, on interaction between a primary and a bound electron with a bond energy great in comparison with the bond energy of conductivity electrons, the angle between the directions of the primary and secondary electrons is appreciably different from 90° and, consequently, in semi-conductors and insulators, with normal incidence of the primary beam on the emitter surface, part of the secondary electrons formed will have a great velocity component in the direction of the surface of the emitter.

However, in the case of the semi-conductors with which we are here concerned, i. e. oxygen-caesium and other emitters, which, as was indicated above, have a large concentration of electrons with low bond energy, the distribution of secondary electron velocities in different directions will be very little different from the corresponding distribution of secondary electron velocities in metals, and thus this circumstance cannot be considered to be a cause of the great secondary emission from composite surfaces.

Results of measurements of the secondary emission from semi-conductors with low concentrations of conductivity electrons, for instance cuprous oxide, provide good confirmation of the suppositions expressed above. The magnitude of the secondary emission from cuprous oxide is less than that from copper. A number of experimental facts quoted below also confirm our suppositions.

It is well-known that the increase in the secondary emission from composite surfaces as compared with the secondary emission from metals is at the expense of secondary electrons with low initial energies, i. e. the energy distribution curves of electrons in metals and in composite surfaces are of quite different shape.

If the mechanism of secondary emission from metals and composite surfaces were the same, as is supposed by Morgulis and Bruining, i. e. if the magnitude of the secondary emission in each of these cases were determined only by the coefficient of absorption of secondary electrons which attain sufficient velocity to overcome the potential barrier when escaping from the emitter, then, as will easily be seen, the energy distribution curves of the secondary electrons should be the same in the two cases. This is confirmed experimentally by the circumstance that the energy distribution curves of secondary electrons are of the same shape for copper and cuprous oxide.

Further, in the event of identical mechanism of secondary emission from conductors and complex emitters, the increase in the exit work in such an emitter should lead to a considerable reduction in the secondary emission of electrons, which is not observed in actuality. For instance, oxidation of an oxygen-caesium emitter, which causes total loss of photoelectric sensitivity and consequently causes a great increase in the escape work, only reduces the secondary emission by 10—20%.

On the basis of a number of other investigations into secondary emission from composite surfaces we also came to the conlusion that the mechanism of secondary emission from such surfaces is essentially different from the mechanism of secondary emission from metals.

In our opinion the great secondary emission from composite surfaces arises as a consequence of the extraction of electrons by positive charges which are formed on the surface of a complex emitter, when it is bombarded by a beam of primary electrons, i. e. to all intents and purposes a Malter effect. The absence of appreciable inertia in secondary emission from composite surfaces, for instance from an oxygen-caesium emitter, does not contradict our assumption. Indeed, in the case under consideration the inertia of secondary emission is determined by the time of recombination of positive charges, which may be less than 10^{-9} sec., whilst each positive charge may nevertheless extract a large number of electrons from the emitter during the course of its existence, since the time necessary for the extraction of an electron from the emitter will be less than 10^{-14} sec. The time of extraction of an electron from the emitter will be determined very approximately, as the time necessary for an electron to travel over a path of 10^{-7} cm with an average velocity of 1 V.

The positive charge which is formed on the surface of an emitter with a low primary current density i_{pr} will be proportional to this value and to the probability of ionization of atoms or molecules which form the surface of the emitter.

As is well-known, the probability of ionization of atoms or molecules depend on the velocity of the primary electrons $Wf(V_{pr})$. If it is assumed that, on an average, each ionized atom or molecule frees K electrons from the emitter and that the probability of their ionization is expressed by the same empirical formula, as was suggested by Compton in calculating the ionization in gases, i. e.:

$$W = C_1 (V - V_i)\, e^{-\frac{V - V_i}{C_2}}$$

where C_1 and C_2 are constants and V_{pr} and V_i are the velocity of the electron and the ionization potential respectively, then for a current i_{sec} the secondary emission is given by:

$$i_{sec} = K i_{pr} C_1 (V - V_i)\, e^{-\frac{V - V_i}{C_2}}. \qquad (1)$$

Supposing that $KC_1 = AC_2 = \frac{1}{\mu}$ and neglecting the magnitude V_i in comparison with V, we arrive at the empirical formula of Lukyanov [9]:

$$i_{sec} = A i_{pr} V e^{-\mu V} \qquad (2)$$

which agrees well with experimental data.

As is well-known, for low velocities of the primary electrons $W = C(V - V_i)$ and consequently:

$$i_{sec} = A i_{pr} (V - V_i) \qquad (3)$$

which is also in good agreement with experiment, since with low velocities of the primary electrons the magnitude of the secondary emission is proportional to the velocity of the primary electrons.

It would seem at first sight that the velocity of the primary electrons, at which the secondary emission attains a maximum, should coincide with the electron velocity corresponding to the maximum ionization of atoms or molecules situated on the surface of the emitter.

However, taking into account that the ionization of atoms or molecules on the surface of the emitter will also take place by way of non-elastic repulsion and owing to secondary electrons which have sufficient velocity for this, it will be understood that the primary electron velocity which corresponds to maximum secondary emission will differ considerably from the electron velocity corresponding to the maximum probability of ionization of the given atoms or molecules. Thus, the large values of the primary electron velocities $V=400$ V, at which the maximum secondary emission is obtained from composite surfaces, cannot serve as an objection to our proposed mechanism of the secondary emission of electrons.

It should be pointed out that in the case of the photoeffect of oxygen-caesium emitters there is also present, in all probability, an effect of extraction of electrons by ions of caesium which are formed on the surface of the emitter by the ionization of adsorbed atoms of caesium. This follows from the experimental data referred to hereafter.

When the surface of an oxygen-caesium emitter is illuminated with infrared light, the volt-ampere curve becomes parallel to the voltage axis at a higher positive potential difference between the emitter and the photoelectron collector, than when the emitter is illuminated by light of short wave-length([7]). In other words, the saturation current is attained with a greater potential difference between the emitter and collector when the emitter is illuminated with long wave-length light than with short.

This is probably because the caesium ions which are formed in the surface film during the illumination of an oxygen-caesium emitter with long wave-length light extract electrons from the emitter. Indeed, in order that the electrons extracted by positive ions should be able to fall onto the collector, it is necessary to create a certain gradient in the external electric field at the cathode, to neutralize the influence of the positive ions on these electrons.

The effect of extraction of electrons from an oxygen-caesium cathode by positive ions also, obviously, explains the great photoelectronic emission from it on illumination with light of long wave-length. It thus becomes comprehensible, why a large photoeffect arises in spite of the small absorption of light by a single-molecular surface film of caesium. This assumption forms a complement to de-Boer's theory.

The effect of interaction between electrons and positive ions formed on the surface of the emitter becomes still more noticeable during the emission of secondary electrons, when it is necessary to raise the potential difference between the emitter and collector to 15—20 V, in order to obtain the saturation current in the absence of the influence of a space-charge([5]).

In this case the increase in the potential difference is due to the circumstance that during the bombardment of the surface of an oxygen-caesium emitter by a beam of primary electrons, ions of compounds of caesium with oxygen are formed on the surface as well as ions of caesium.

Cracks are formed in the intermediate layer between the colloidal particles in complex emitters of the oxygen-caesium type, and there are consequently set up potential barriers which reflect the slow secondary electrons well. When the escape-work of electrons from the surface of a complex emitter

is small, the reflection of electrons from these barriers may facilitate the large secondary emission, which probably explains the fact that oxidation of the surface of an oxygen-caesium emitter, which causes total loss of its photosensitivity, reduces the coefficient of secondary emission δ by two or three units.

Assuming that the large secondary emission from complex emitters is mainly a result of the extraction of electrons from the emitter by positive charges which are formed on its surface during bombardment of it by a beam of primary electrons, we come to the following conclusion.

The magnitude of the secondary emission from composite surfaces is mainly determined by the time of recombination of positive ions T and by the magnitude K which characterizes the average number of electrons which are extracted by unit charge in unit time.

It is quite clear that the magnitudes K and T are determined by the electrical conductivity of the intermediate layer. It follows immediately from this that secondary emission from complex emitters depends both on the nature and on the relative quantity of particles of metals and their compounds which enter into the composition of the intermediate layer.

We have indicated these facts several times in previous works.

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TECHNICAL PHYSICS

SOUND ENERGY DENSITY IN AN ENCLOSURE EXCITED BY DIRECTIONAL SOURCE

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In a number of cases it is convenient to consider the sound energy in a given enclosure as consisting of two parts—the energy due to direct sound and the energy due to reflected sound. Accordingly, the sound energy density may be represented by the sum

$$E = E_d + E_r \tag{1}$$

where E is total sound energy density at a certain point of the enclosure, while E_d and E_r are its parts due to the direct radiation and reflection respectively.

From the standpoint of the so-called statistical theory, value E is the sum of terms of an infinite series, in which the n-th term equals the total energy of all reflections of the n-th order. Thus

$$E = \frac{W r_m}{cV}(1 + \beta_{av} + \beta_{av}^2 + \beta_{av}^3 + \ldots) \tag{2}$$

where r_m is mean path between two successive reflections, c is sound velocity, V is volume of the enclosure, $\beta_{av} = 1 - \alpha_{av}$ is average coefficient of reflection, $\alpha_{av} = \frac{\sum \alpha S}{\sum S} = \frac{A}{\sum S}$ is average coefficient of absorption, A is total absorption, $\sum S$ is total surface area of the enclosure.

For a rectangular enclosure

$$r_m = \frac{4V}{\sum S}. \tag{3}$$

After substitution into (2) and summation we get a well-known expression for the total energy density

$$E = \frac{4W}{c(1 - \beta_{av})\sum S} = \frac{4W}{Ac}. \tag{4}$$

The reflected energy density is obtained by discarding from the right part in (2) the member proportional to the direct sound energy, i. e. the first term of the series, which is equal to unity

$$E_r = \frac{W r_m}{cV}(\beta_{av} + \beta_{av}^2 + \beta_{av}^3 + \ldots) \tag{5}$$

or after transformations

$$E_r = \frac{4W}{Ac}(1 - \alpha_{av}). \tag{6}$$

Making good use of the conception of imaginary sources arising by reflection, we can present the reflected energy density as the total density, provided that there is not one source of sound with power W, but a series of sources with total power

$$W_1 = W\beta_{av} = W(1 - \alpha_{av}).$$

Each of them is an imaginary source of the first reflection, corresponding to one of the enclosure's surfaces. By substituting W_1 for W in (4), the expression

$$E_r = \frac{4W_1}{Ac} \tag{7}$$

analogous to (6) is obtained.

There are three cases in which the energy density expressions (4) and (6) may be applied: 1) non-directional source in an enclosure with uniformly distributed absorption; 2) non-directional source in an enclosure with non-uniformly distributed absoption; 3) directional source in an enclosure with uniformly distributed absorption.

In the fourth case, which is that of a directional source and a non-uniformly distributed absorption, the expressions fail to provide accurate results. Indeed, let us imagine a room with a very high absorption by one of its surfaces and a low absorption by the other surfaces, and a directional source to be placed in it whose radiation concentrated into a narrow beam is toward the highly absorbing surface. Under these conditions a predominant part of sound energy will be lost in the first reflection, and the reflected sound energy density will be in general small. If, however, the radiation of the same source is toward a weakly absorbing surface, reflections of high order are required for such a loss to occur, and its influence on the total energy balance will be comparatively small. It will then be obvious that the reflected sound energy density may in that case be changed by altering merely the orientation of the source and this is not consistent with the statistical theory.

The latter does not hold here, owing to the fact that with a small number of reflections of the first order, each of them has a substantial influence on the total. In reality we have the number of reflections of the n-th order

$$N = (2n)^2 + 2 \tag{8}$$

i. e. 6 first reflections are followed by 18 second reflections, 38 third reflections, and so on.

The expression (6) may be improved, if the reflections of the first order are isolated and taken into account separately. Then, as before, in considering the reflected sound energy density as the total energy density from the imaginary sources of the first order, we have only to determine their total power W_1.

Of great practical interest is the case of an auditorium or concert-hall provided with sound reinforcing system. Here we have a highly absorbing rectangular area (audience) of one of the surfaces of the enclosure (floor) and weakly absorbing other surfaces, while the directional source of sound (loudspeaker) is radiating toward the highly absorbing surface.

In order to calculate the power W_1 it is practical to isolate the part of power W falling on the area occupied by the highly absorbing audience. It may be written

$$W_{\square} = WQ \tag{9}$$

where W is power of the loudspeaker, Q is the relative value of power, falling on the highly absorbing surface.

Then, assuming the absorption to be uniform over the remaining surfaces of the enclosure, we can write

$$W_1 = W[Q(1-\alpha_1)+(1-Q)(1-\alpha_2)] \quad (10)$$

where α_1 is a coefficient of absorption of the highly absorbing surface, α_2 is average coefficient of absorption of the remaining surfaces of the enclosure.

By substituting W_1 from (10) into (7) we obtain:

$$E_r = \frac{4W}{Ac}[Q(1-\alpha_1)+(1-Q)(1-\alpha_2)]. \quad (11)$$

Now, let us determine the value of Q. For this purpose we turn to the schematic drawing (Fig. 1), in which a highly absorbing area having the sides (x_2-x_1) and $2y_2$ is in the vicinity of a loudspeaker suspended at a high h and whose axis forms an angle α with the vertical and lies in a plane, which is perpendicular to the side $2y_2$ of the considered area and divides the latter into two equal parts. Intercept L is assumed to be equal to unity.

Let us calculate the energy falling on the elementary area with sides Δx and Δy. The elementary solid angle, with its summit in the point of suspension of the loudspeaker, and to be based on the elementary area, is

$$\Delta\Omega = \frac{\Delta x \Delta y \cos\vartheta}{r^2} = \frac{\Delta x \Delta y \operatorname{ctg}\alpha}{r^3}. \quad (12)$$

Further taking into consideration that $r = (x^2+y^2+\operatorname{ctg}^2\alpha)^{\frac{1}{2}}$, we have:

$$\Delta\Omega = \frac{\Delta x \Delta y \operatorname{ctg}\alpha}{(x^2+y^2+\operatorname{ctg}^2\alpha)^{\frac{3}{2}}}; \quad (13)$$

The energy radiated within the elementary angle Ω is

$$\Delta W = \frac{W}{4\pi} n_\vartheta \Delta\Omega \quad (14)$$

where n_ϑ is concentration coefficient of the loudspeaker in the direction ϑ. The concentration coefficient in the direction ϑ may be expressed through the axial-concentration coefficient and the directional coefficient as follows *:

$$n_\vartheta = n_0 R_\vartheta^2 \quad (15)$$

where n_0 is the axial-concentration coefficient, R_ϑ is the directional coefficient at angle ϑ to the source axis.

Furthermore ([1])

$$n_0 = \frac{2}{\int_0^\pi R_\vartheta^2 \sin\vartheta d\vartheta}. \quad (16)$$

* The concentration coefficient is the ratio of the sound intensity in a given direction to the average intensity for all directions measured at the same distance from the source. The directional coefficient is the ratio of sound pressure in a given direction to the sound pressure on the axis at the same distance from the source.

For calculation of the energy $W_{\square}$ falling on the surface having the sides $(x_2 - x_1)$ and $2y_2$ we must integrate ΔW over the whole surface

$$W_{\square} = 2 \int_{x_1}^{x_2} \int_{0}^{y_2} \frac{W}{4\pi} n_0 R_\vartheta^2 \frac{\operatorname{ctg} \alpha \, dx \, dy}{(x^2 + y^2 + \operatorname{ctg}^2 \alpha)^{\frac{3}{2}}}. \tag{17}$$

Taking the constant coefficients outside of the integral and dividing the right and left parts by W, we obtain from (9)

$$Q = \frac{n_0 \operatorname{ctg} \alpha}{2\pi} \int_{x_1}^{x_2} \int_{0}^{y_2} \frac{R_\vartheta^2 \, dx \, dy}{(x^2 + y^2 + \operatorname{ctg}^2 \alpha)^{\frac{3}{2}}}. \tag{18}$$

The value R_ϑ^2 remains under the integration sign, because ϑ is a function of x and y. It may be shown that

$$\cos \vartheta = \frac{1 - (1 - x) \sin^2 \alpha}{(x^2 + y^2 + \operatorname{ctg}^2 \alpha)^{\frac{1}{2}} \sin \alpha}. \tag{19}$$

We will now proceed to calculate the coefficient Q for real directional characteristics. The horn loudspeakers in general use have, as shown by our investigations [1], directional characteristics in the form of ellipsoid of revolution*. With such a form of directional characteristic, the directional coefficient is

$$R_\vartheta = \frac{(1 - e^2) \cos \vartheta}{1 - e^2 \cos \vartheta} \tag{20}$$

where e is eccentricity of the ellipse equivalent to the directional characteristic. The axial-concentration coefficient

$$n_0 = \frac{4e^2}{(1 - e^2)^2 \left[\frac{1}{1 - e^2} + \frac{1}{2e} \ln \frac{1 + e}{1 - e} \right]}. \tag{21}$$

Placing (20) and (21) into (17) and taking into account (19), we obtain after transformation

$$Q = \frac{n_0 (1 - e^2)^2 \operatorname{ctg} \alpha}{2\pi \sin^2 \alpha} \times$$

$$\times \int_{x_1}^{x_2} \int_{0}^{y_2} \frac{[1 - \sin^2 \alpha (1 - x)]^2 \, dx \, dy}{[y^2 + (x^2 + \operatorname{ctg}^2 \alpha)]^{\frac{1}{2}} \left\{ y^2 + (x^2 + \operatorname{ctg}^2 \alpha) - \frac{e^2}{\sin^2 \alpha} [1 - \sin^2 \alpha (1 - x)]^2 \right\}^2}. \tag{22}$$

The expression (22) may be reduced to the form

$$Q = \frac{n_0 (1 - e^2)^2 \operatorname{ctg} \alpha}{2\pi \sin^2 \alpha} \int_{x_1}^{x_2} \int_{0}^{y_2} \frac{b^2 dx \, dy}{(y^2 + a^2)^{\frac{1}{2}} (y^2 + d^2)^2} \tag{23}$$

where a, b, d are functions of x.

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* On the assumption that the sound source is located at the end of the ellipse's major axis having its own axis coinciding with the latter.

Integration with respect to y and substitution of limiting values gives

$$Q = \frac{n_0 (1-e^2)^2 \operatorname{ctg} \alpha}{2\pi \sin^2 \alpha} \int_{x_1}^{x_2} \frac{b^2}{2d^2 (a^2-d^2)} \left[\frac{y_2 \sqrt{y_2^2+a^2}}{y_2^2+d^2} + \right.$$

$$\left. + \frac{a^2-2d^2}{d\sqrt{a^2-d^2}} \operatorname{arctg} \frac{y_2 \sqrt{a^2-d^2}}{d\sqrt{y_2^2+a^2}} \right] dx. \qquad (24)$$

To determine the integral in (24) analytically is of considerable difficulty, and we have therefore resorted to the graphical method. In accordance with the experimental data ([1]), we have taken $e=0.9$ and $n_0=25$, what is true for sufficiently large frequency range.

Further, we have taken five values for y_2 ranging within $0.125-0.75$ and $\alpha=80°$. The curves in Fig. 2 represent the function $\frac{dQ}{dx}=f(x)$ in this case.

Determining the value of Q from one of the curves shown in Fig. 2 is accomplished by measuring * the area enclosed by the curve within

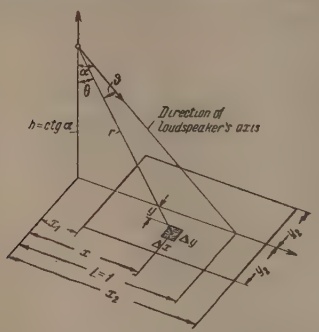

Fig. 1. Relative position of a loudspeaker and a highly absorbing surface.

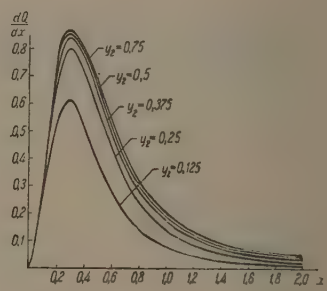

Fig. 2. Dependence $\frac{dQ}{dx}=f(x)$ for various y_2 at $\alpha=80°$.

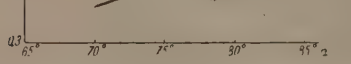

Fig. 3. Dependence $Q=f(\alpha)$ at $x_1=0$ and $x_2=1$ for $y_2=0.5$ and 0.25.

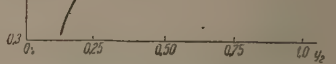

Fig. 4. Dependence $Q=f(y_2)$ at $x_1=0$ and $x_2=1$ for $\alpha=80°$.

the required limits of x. The dependence of Q thus determined within the range $x_1=0$, $x_2=1$ ** upon the angle α at $y_2=0.5$ and 0.25 is shown in Fig. 3, while in Fig. 4 is shown the dependence of Q upon y_2 at $\alpha=80°$. From Figs. 3 and 4 it is seen that Q does not exceed 0.5, and this was to be expected indeed, because the whole upper half of the sound beam radiated by the loudspeaker falls outside of the highly absorbing surface. In that case only the value of Q can exceed 0.5,

* On the scale of the co-ordinate system.

** With $x_2=1$ the axis of the loudspeaker is directed toward the far side of the absorbing surface.

when the axis of the loudspeaker is directed toward the central part of the absorbing surface, i. e. when $x_2 > 1$. It should be noted, however, that with x_2 increase from 1 to 2, Q does nos change appreciably while the direct sound energy distribution over the surface in question becomes far less uniform.

Figs. 3 and 4 show also that if $y_2 > 0.4$ and $\alpha > 70°$, at $x_1 = 0$ and $x_2 = 1$, Q may be considered practically equal to 0.5.

The method of determining Q described here applies to the case of a loudspeaker located in the plane of symmetry of the absorbing surface. With a loudspeaker placed asymmetrically the problem should be reduced to two symmetrical cases (by completing the surface accordingly). The value of Q for the given surfaces is thus determined as the half difference between the values of Q for two symmetrical surfaces, or half sum of these values.

Since the reflected sound energy is uniformly distributed in the enclosure, in general, expression (11) holds true for any point of the enclosure. Now let us deal with the direct sound energy from the loudspeaker, and its distribution over the considered surface. We will take advantage of the data of our other work [2] and offer an expression for the sound intensity in an arbitrary point of the enclosure. This expression for ellipsoidal directional characteristic is as follows:

$$I_{x,y} = \frac{(1-e^2)^2 \cos^2(\alpha-\varphi)\sin^2\varphi\cos^4\psi}{[1-e^2\cos^2(\alpha-\varphi)\cos^2\psi]^2 x^2}\frac{Wn_0}{4\pi} = g\frac{Wn_0}{4\pi} \qquad (25)$$

where x and y are co-ordinates of the point given *, $\varphi = \operatorname{arctg}\frac{x}{h}$, $\psi = \operatorname{arctg}\frac{y\cos\varphi}{h}$, and the remaining known from the preceding value.

The so-called acoustical ratio is determined by the following expression:

$$\eta_l = \frac{E_r}{E_d}. \qquad (26)$$

By using (25) and also (11), and taken into account that $I = Ec$, we can state that the acoustical ratio in an arbitrary point of the surface is:

$$\eta_{x,y} = \frac{16\pi[Q(1-\alpha_1)+(1-Q)(1-\alpha_2)]}{Ag\,n_0}. \qquad (27)$$

And finally the addition to the sound intensity level (in decibels) at an arbitrary point (with respect to the direct sound) due to the reflected sound energy is:

$$\mu_{x,y} = 10\lg(\eta_{x,y}+1). \qquad (28)$$

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* Here the co-ordinates must be expressed in absolute units.

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TECHNISCHE PHYSIK

ÜBER EINIGE THEOREME DER STÖRUNGSTHEORIE UND INSBESONDERE ÜBER DEREN ANWENDUNG IN DER THEORIE DER NICHTSTATIONÄREN ERSCHEINUNGEN IN ELEKTRONENRÖHREN

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(Vorgelegt von N. D. Papalexi, Mitglied der Akademie, d. 17. VI. 1939)

1. In der Störungstheorie begegnet man zweierlei Problemen, von denen die einen die Bestimmung der gestörten Bewegung für unendilch lange Zeiten erfordern, während bei den anderen nur gewisse endliche Zeitintervalle in Betracht kommen. Zu den ersten gehören, z. B., die die Stabilität der Bewegung betreffenden Fragen, zu den zweiten—sehr viele Fragen der technischen Physik, von denen wir hier nur die Theorie der rasch veränderlichen Prozesse in Elektronenröhren, in denen die Elektronen der Einwirkung des Feldes nur während ihrer Durchlaufszeit durch die Anordnung ausgesetzt sind, erwähnen. Besteht dabei die Spannung, unter deren Wirkung die Bewegung erfolgt, aus einem konstanten Teile, dem eine im Vergleich zu ihm kleine veränderliche Zusatzspannung überlagert ist, so hat man es gerade mit einer Aufgabe der zweiten Art zu tun. Dabei stimmen auch die Anfangsbedingungen für die ungestörte und die gestörte Bewegung (U. B. und G. B.) überein. Man kann deshalb für die Koordinaten (verallgemeinerten) der G. B. den Ansatz $x_i = x_i^0 + u_i$ machen, wobei x_i^0 der U. B. entspricht und u_i bei genügender Kleinheit der Störungskräfte während der ganzen Bewegungszeit (endlichen) beliebig klein bleibt.

Wir geben hier einige Theoreme, die die Lösung derartiger Fragen ermöglichen, wobei, obwohl sie eigentlich zur Behandlung der Probleme zweiter Art unmittelbar bestimmt sind, sie jedoch auch mit geringen Abänderungen auf Probleme der ersten Gruppe anwendbar sind. Wir haben die Absicht auf diese und verwandte Fragen an anderem Orte näher einzugehen, wobei die Methode näher erleutert und durch eine grössere Zahl von Beispielen illustriert werden soll.

2. a) Wir beginnen mit dem einfachsten Falle eines Systems mit nur einem Freiheitsgrade, das durch die Koordinate ξ charakterisiert wird. Die Bewegungsgleichung und die Anfangsbedingungen lauten:

$$\ddot{\xi} = f(\xi, \dot{\xi}) + \varepsilon\psi(\xi, \dot{\xi}, t), \quad (\xi)_{t=0} = \xi_0, \quad (\dot{\xi})_{t=0} = \dot{\xi}_0, \tag{1}$$

wobei f von der Zeit explicite nicht abhängt und ε — ein kleiner Parameter ist. Wir setzen $\xi = \xi^0 + u$, wo ξ^0 der U. B. entspricht, also

$$\ddot{\xi}^0 = f(\xi^0, \dot{\xi}^0), \quad (\xi^0)_{t=0} = \xi_0, \quad (\dot{\xi}^0)_{t=0} = \dot{\xi}_0, \tag{2}$$

und erhalten als Gleichung für die erste Näherung von u:

$$\ddot{u} = \left(\frac{\partial f}{\partial \dot{\xi}}\right)_0 \dot{u} + \left(\frac{\partial f}{\partial \xi}\right)_0 u + \varepsilon\psi(\xi^0, \dot{\xi}^0, t),\ (u)_{t=0}=0,\ (\dot{u})_{t=0}=0. \qquad (3)$$

Dies ist eine inhomogene lineare Gleichung mit variablen Koeffizienten. Indessen lässt sie sich sofort durch Quadraturen integrieren, falls die Lösung der Gl. (2) der U. B. bekannt ist, da die Grösse $v=\dot{\xi}^0$, wie man sofort einsieht, ein partikulares Integral der entsprechenden homogenen Gleichung, d. h.

$$\ddot{v} = \left(\frac{\partial f}{\partial \dot{\xi}}\right)_0 \dot{v} + \left(\frac{\partial f}{\partial \xi}\right)_0 v \qquad (A)$$

ist.

Dieser Umstand, von dessen Richtigkeit man sich übrigens unmittelbar durch Differentiation der Gl. (2) überzeugt, ist damit verbunden, dass diese Gleichung, in der die Zeit explicite nicht auftritt, eine einparametrige Transformationsgruppe zulässt, die ihre Form unverändert lässt. Ist nähmlich $\xi^0=\varphi(t)$ eine Lösung von (2), so ist auch $\xi_1=\varphi(t+\alpha)$, wo α — eine beliebige Konstante bedeutet, wiederum eine Lösung. Indem man α unendlich klein voraussetzt und nach Potenzen von α entwickelt, ersieht man sofort, dass $\left(\frac{\partial\varphi}{\partial t}\right)_{\alpha=0}=\dot{\xi}^0$ eine Lösung von (A) sein muss.

Insofern eine particulare Lösung $v_1=\dot{\xi}^0$ von (A) bekannt ist, erhält man sofort auch deren allgemeines Integral. Bedeutet v_2 irgend eine zweite partikulare Lösung von (A), und ist Δ die aus v_1 und v_2 gebildete Determinante, also

$$\Delta=\Delta(t)=v_2\dot{v}_1-\dot{v}_2v_1=\Delta(0)\cdot\exp\left\{\int_0^t\left(\frac{\partial f}{\partial\dot{\xi}}\right)_0 d\tau\right\}, \qquad (4)$$

so lässt sich die Lösung von (3) in folgender Form schreiben:

$$u=\varepsilon\int_0^t [v_1(t)\,v_2(\tau)-v_2(t)\,v_1(\tau)]\,\frac{\chi(\tau)}{\Delta(\tau)}\,d\tau, \qquad (5)$$

wobei $\chi(\tau)=\{\psi(\xi^0,\dot{\xi}^0,t)\}_{t=\tau}$ ist.

b) Die bisherige Betrachtung bezog sich nur auf die erste Näherung. Es ist indessen leicht zu ersehen, dass falls die Funktionen $f(\xi,\dot{\xi})$ und $\psi(\xi,\dot{\xi},t)$ die Bedingungen erfüllen, unter denen die Lösung u eine holomorphe Funktion von ε ist (1), so lassen sich auch die höheren Näherungen auf Quadraturen zurückführen. In der Tat, setzt man u in Form einer Potenzreihe an, also $u=\sum_{k=1}^{\infty}u_k\varepsilon^k$, und trägt man $\xi=\xi^0+u=\xi^0+\sum_{(k)}u_k\,\varepsilon^k$ in (1) ein, so findet man durch Nullsetzen der Koeffiziente bei jeder Potenz von ε:

$$\ddot{u}_1=\left(\frac{\partial f}{\partial\dot{\xi}}\right)_0\dot{u}_1+\left(\frac{\partial f}{\partial\xi}\right)_0 u_1+\psi(\xi^0,\dot{\xi}^0,t),\ldots,\ddot{u}_k=$$

$$=\left(\frac{\partial f}{\partial\dot{\xi}}\right)_0\dot{u}_k+\left(\frac{\partial f}{\partial\xi}\right)_0 u_k+F_k,\ldots\quad(k=1,\,2,\ldots\infty), \qquad (6)$$

wobei in F_k nur Funktionen u_s mit $s<k$ auftreten. Man erhält also für alle u_k ein und dieselbe lineare Gleichung, wobei sich nur die rechten Seiten der verschiedenen Gleichungen voneinander unterscheiden. Ihre Integration geschieht deshalb in derselben Weise wie die von (3), also in Quadraturen.

Beispiel. Die Bewegungsgleichung eines Elektrons in einer zylindrischen Diode unter der Einwirkung einer konstanten Anodenspannung φ_0 hat die Form $\ddot{\xi} = 1/\xi^0$ und lässt sich mittels Quadraturen integrieren. Wirkt dabei noch eine verhältnismässig kleine veränderliche Zusatzspannung, so wird die Bewegungsgleichung zu $\ddot{\xi} = [1 + \varepsilon\psi(t)]/\xi$, wobei $|\varepsilon\psi(t)| \ll 1$ ist. Gl.(A) für v wird jetzt $\ddot{v} = -v/(\xi^0)^2$, und da ihr $v = \dot{\xi}^0$ genügt, so lässt sich die G. B. sofort mittels Quadraturen bestimmen.

3. Liegt die allgemeinere Gleichung

$$\ddot{\xi} = f(\xi, \dot{\xi}, t) + \varepsilon\psi(\xi, \dot{\xi}, t) \tag{7}$$

vor, wo f von der Zeit explicite abhängt, so lässt sich die Integration der Gleichungen für die Störungen u_k verschiedener Ordnungen, die wieder von der Form (6) sind, vollziehen, falls eine einparametrige Transformationsgruppe bekannt ist, die die Form der Gleichung der U. B., also

$$\ddot{\xi} = f(\xi, \dot{\xi}, t) \tag{B}$$

ungeändert lässt. Denn ist α der Parameter der zu betrachtenden Gruppe und $\xi = \varphi(t, \alpha)$ ein Integral von (B), das für $\alpha = \alpha_0$ den erforderlichen Anfangsbedingungen entspricht, so ist offenbar $v = \left(\frac{\partial \varphi}{\partial \alpha}\right)_{\alpha=\alpha_0}$ eine partikulare Lösung der homogenen Gleichung, die der G. B. entspricht, also $\ddot{v} = \left(\frac{\partial f}{\partial \dot{\xi}}\right)_0 \dot{v} + \left(\frac{\partial f}{\partial \xi}\right)_0 v$. Dadurch lässt sich jede der linearen Gleichungen (6) in eine lineare Gleichung erster Ordnung verwandeln, die sich sofort integrieren lässt.

Beispiel. Die Gl. $\ddot{\xi} = [t + \varepsilon\psi(t)]/\xi^m$, $m = \text{const}$, bestimmt die Bewegung der Elektronen in einer zylindrischen ($m = 1$) oder sphärischen ($m = 2$) Diode, unter gewissen Bedingungen. Die Gleichung der U. B. $\ddot{\xi}^0 = t/(\xi^0)^m$ bleibt bei der Transformation $\xi = \alpha\xi'$, $t = \alpha^{\frac{m+1}{3}} t'$ unverändert. Ist also $\xi^0 = \varphi(t)$ eine Lösung der Gleichung der U. B., die den erforderlichen Anfangsbedingungen genügt, so ist $\xi' = \varphi(t')$, also $\xi = \alpha\varphi\left[t\alpha^{-\frac{m+1}{3}}\right]$ — auch eine Lösung, die für $\alpha = \alpha_0 = 1$ in die ursprüngliche übergeht. Es ist also $v = \left[\frac{\partial \xi}{\partial \alpha}\right]_{\alpha=1} = \varphi(t) - \frac{m+1}{3} t\varphi'(t) = \xi^0 - \frac{m+1}{3} t\dot{\xi}^0$ eine Losüng der homogenen Gleichung, die der G. B. entspricht, und dadurch gelangt man zur gesuchten Lösung (²).

Wir betonen hier nochmals, dass es sich in § 2 und § 3 um die Ausnützung von partikularen Lösungen der Gleichung der U. B. zur Bestimmung der G. B. handelt. Da diese Lösungen ganz bestimmten Anfangsbedingungen entsprechen, also keine willkürlichen Konstanten enthalten, so können hier auch z. B. durch numerische Integration gefundene Lösungen benutzt werden.

4. Die Theoreme von § 2 und § 3 lassen sich unmittelbar auf Systeme mit beliebiger Anzahl von Freiheitsgraden verallgemeinern, worauf wir hier nicht näher eingehen können.

5. Ist nicht nur eine partikulare Lösung der Gleichungen der U. B. bekannt, sondern auch irgend ein willkürliche Konstanten enthaltendes Integral, so lässt es sich natürlicherweise auch zur Bestimmung der G. B. benutzen, insbesondere im Verein mit der Methode der §§ 2 und 3. Schreibt man, nämlich, die Bewegungsgleichungen in der Form $x_k =$

$= f_k(x_1, x_2, \ldots, x_n, t) + \varepsilon\psi_k(x_1, x_2, \ldots, x_n, t)$, $k = 1, 2, \ldots n$, und kennt man ein Integral der U. B. — $F(x_1, x_2, \ldots, x_n, t) = C$, dass keine anderen willkürlichen Konstanten ausser C enthält, so wird $\frac{dF}{dt} = \sum_{k=1}^{(n)} \frac{\partial F}{\partial x_k}\dot{x}_k + \frac{\partial F}{\partial t}$ identisch Null, falls man dahin die $\dot{x}_k$ aus den Bewegungsgleichungen einträgt, wobei man dort $\varepsilon = 0$ zu setzen hat. Im Falle der G. B. erhält man mittels derselben Formeln, aber für $\varepsilon \neq 0$:

$$F = F(x_1, x_2, \ldots, x_n, t) = C + \varepsilon \sum_{k=1}^{k=n} \int_0^t \psi_k(x_1, x_2, \ldots, x_k, t)\, dt. \qquad (8)$$

Beschränkt man sich auf die Betrachtung der ersten Näherung (die Erweiterung der Resultate auf höhere Näherungen bietet keine Schwierigkeiten), so folgt daraus mit Rücksicht darauf, dass die Anfangsbedingungen für die U. B. und die G. B. gleich sind:

$$F = F(x_1, x_2, \ldots, x_n, t) = C + \varepsilon \sum_{k=1}^{k=n} \int_0^t \psi_k(x_1^0, x_2^0, \ldots, x_n^0, t)\, dt.$$

Unter Benutzung solcher Beziehungen für jedes bekannte Integral der U. B. kann man einige Veränderlichen eliminieren und so in manchen Fällen zur Möglichkeit der Ausnützung der Resultate von § 2 und § 3 gelangen, falls man sogar mit den vorhandenen Integralen für die vollständige Lösung des Problems nur mit deren Hilfe allein nicht auskommt (man vgl. das Beispiel des russichen Textes und § 6).

6. Zum Schluss sei noch Folgendes bemerkt: sind für die Bewegungsgleichungen des § 5 n unabhängige Integrale vom dort betrachteten Typus, d. h. $F_i(x_1, x_2, \ldots, x_n, t) = C_i$, $i = 1, 2, \ldots, n$, bekannt, so erhält man, indem man für sie die Beziehungen (8) hinschreibt und deren linke Seiten nach Potenzen der u_k entwickelt, n Gleichungen von der Form:

$$\sum_{k=1}^{k=n} \left(\frac{\partial F_i}{\partial x_k}\right)_0 u_k = \varepsilon \sum_{k=1}^{k=n} \int_0^t \psi_k(x_1^0, x_2^0, \ldots, x_n^0, t)\, dt,$$

$$i = 1, 2, \ldots, n, \qquad (x_k = x_k^0 + u_k),$$

wobei nur die erste Näherung in Betracht gezogen ist. Daraus lassen sich sofort alle u_k bestimmen. Ähnliche Überlegungen lassen sich auch zur Bestimmung höherer Näherungen verwenden, die in analoger Weise gefunden werden können (siehe 2b).

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ZITIERTE LITERATUR

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PHYSIQUE DU GLOBE

ÉTUDES DES VARIATIONS SÉCULAIRES DU MAGNÉTISME TERRESTRE. I. VARIATIONS SÉCULAIRES DÉDUITES DE LA THÉORIE DE BIOT

Par P. P. LASAREFF (P. P. LAZAREV), de l'Académie

§ 1. Biot(1) a démontré que le champ magnétique terrestre peut être approximativement représenté par celui d'un aimant infiniment petit—un dipôle, placé au centre du globe avec l'axe magnétique incliné sous un angle de 11° environ à l'axe de rotation de la terre. Les travaux géniaux de Gauss(2) ont permis de trouver l'expression précise du potentiel magnétique du globe pour tout espace qui l'entoure. Il suit de ces recherches que la solution de Biot correspond à la première approximation du champ réel terrestre. Si le dipôle de Biot reste, par rapport au globe, immobile, les éléments de géomagnétisme ont à chaque point de la surface de la terre les valeurs constantes, indépendantes du temps. Les observations montrent au contraire que les éléments géomagnétiques varient avec le temps d'observation et ces variations lentes portent le nom des variations séculaires.

Il est naturel d'admettre qu'avec une certaine approximation les variations séculaires peuvent être représentées mathématiquement, si l'on admet que l'axe de dipôle de Biot passe toujours par le centre du globe et décrit un cône circulaire.

§ 2. Nous voulons admettre que le centre du dipôle de Biot est placé au centre de la terre O (Fig. 1). Les pôles $+m$ et $-m$ du dipôle ont la distance L. Le potentiel au point A de la surface de la terre est égal à

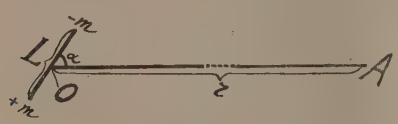

Fig. 1.

$$V = -\frac{mL}{r^2}\cos\alpha = -\frac{\omega}{r^2}\cos\alpha; \quad (1)$$

r est la distance OA, qui est infiniment grande par rapport à L; α est l'angle entre l'axe du dipôle $(+m)O(-m)$ et la droite OA; ω est le moment magnétique du dipôle, égal à celui du globe.

De la formule (1) on déduit les valeurs de la composante verticale Z, de la composante horizontale H, de l'inclinaison J et de la force totale F du champ terrestre.

$$Z = -\frac{\partial V}{\partial r} = -\frac{2\omega}{r^3}\cos\alpha; \quad H = +\frac{1}{r}\frac{\partial V}{\partial \alpha} = \frac{\omega}{r^3}\sin\alpha \quad (2)$$

$$F = \sqrt{Z^2 + H^2} = \frac{\omega}{r^3}\sqrt{1 + 3\cos^2\alpha} \quad (3)$$

$$\operatorname{tg} J = \frac{Z}{H} = -2\operatorname{ctg}\alpha. \quad (4)$$

§ 3. Z, H, F et J [voir formules (2), (3), (4)] sont exprimées en fonction de l'angle α. Nous allons introduire au lieu de α les coordonnées géographiques du point A [latitude φ et longitude O (Fig. 2)] et du pôle nord magnétique n (latitude ϑ et longitude λ). Oz est l'axe de rotation de la terre et On est l'axe magnétique du globe.

Exprimons l^2 (Fig. 2) par deux voies différentes. Le triangle AKn donne

$$l^2 = f^2 + d^2 = (b-a)^2 + d^2. \qquad (5)$$

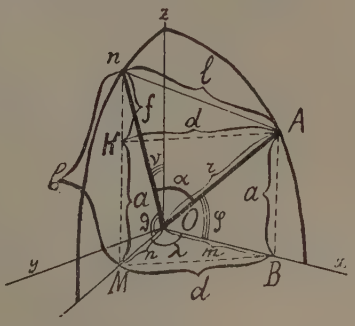

Fig. 2.

Nous avons ensuite

$$d^2 = m^2 + n^2 - 2mn \cdot \cos\lambda$$

avec

$$m = r\cos\varphi;\quad n = r\cos\vartheta;$$
$$a = r\sin\varphi;\quad b = r\sin\vartheta.$$

Après les substitutions correspondantes nous obtenons

$$l^2 = r^2\{(\sin\vartheta - \sin\varphi)^2 + \cos^2\vartheta + \cos^2\varphi - 2\cos\vartheta\cos\varphi\cos\lambda\}$$

d'où nous avons

$$l^2 = 2r^2\{1 - (\sin\vartheta\sin\varphi + \cos\vartheta\cos\varphi\cos\lambda)\}.$$

Le triangle AOn permet de calculer l^2 en fonction de α. Nous avons

$$l^2 = 2r^2(1 - \cos\alpha).$$

En égalant deux expressions de l^2, on a

$$\cos\alpha = \sin\vartheta\sin\varphi + \cos\vartheta\cos\varphi\cos\lambda \qquad (6A)$$

d'où l'on reçoit

$$\sin\alpha = \sqrt{1 - (\sin\vartheta\sin\varphi + \cos\vartheta\cos\varphi\cos\lambda)^2}. \qquad (6B)$$

Les formules (6A) et (6B) expriment $\cos\alpha$ et $\sin\alpha$ en fonction des coordonnées du point A (latitude φ, longitude O) et celles du pôle nord magnétique (latitude ϑ, longitude λ). Après la substitution des valeurs $\cos\alpha$ et $\sin\alpha$ [(6A) et (6B)] dans les équations (2) et (4) nous recevons

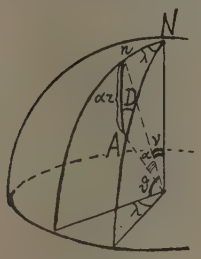

Fig. 3.

$$\left.\begin{aligned} Z &= -\frac{2\omega}{r^3}(\sin\vartheta\sin\varphi + \cos\vartheta\cos\varphi\cos\lambda) \\ H &= \frac{\omega}{r^3}\sqrt{1 - (\sin\vartheta\sin\varphi + \cos\vartheta\cos\varphi\cos\lambda)^2} \\ \operatorname{tg} J &= -2\,\frac{\sin\vartheta\sin\varphi + \cos\vartheta\cos\varphi\cos\lambda}{\sqrt{1 - (\sin\vartheta\sin\varphi + \cos\vartheta\cos\varphi\cos\lambda)^2}} \end{aligned}\right\} \qquad (7)$$

§ 4. Nous allons exprimer la déclinaison D en fonction des coordonnées géographiques du point A et du pôle nord magnétique n (Fig. 3). N est le pôle nord géographique. La déclinaison au point A est égale à l'angle $nAN = D$. Le triangle sphérique nAN donne

$$\frac{\sin D}{\sin\lambda} = \frac{\sin\nu}{\sin\alpha}.$$

En substituant $\nu = \frac{\pi}{2} - \vartheta$, nous avons

$$\frac{\sin D}{\sin\lambda} = \frac{\cos\vartheta}{\sin\alpha} \quad \text{ou} \quad \sin D = \frac{\cos\vartheta\sin\lambda}{\sin\alpha}.$$

Après la substitution de la valeur de sin α [form. (6B)] nous avons

$$\sin D = \frac{\cos\vartheta \sin\lambda}{\sqrt{1-(\sin\vartheta\sin\varphi+\cos\vartheta\cos\varphi\cos\lambda)^2}}. \tag{8}$$

§ 5. Dans les formules (7) et (8) la latitude φ du point A d'observation et sa longitude $\lambda_1 = 0$ sont constantes. La latitude ϑ du pôle nord magnétique, sa longitude λ et le moment magnétique du globe ω sont en général des fonctions de temps t. Nous pouvons donc écrire

$$\vartheta = \vartheta(t);\ \lambda = \lambda(t) \text{ et } \omega = \omega(t).$$

Par un procédé de tâtonnement on peut choisir des fonctions $\vartheta(t)$, $\lambda(t)$, $\omega(t)$ de telle façon que les valeurs Z, H, D, J calculées d'après les formules (7) et (8) coïncident avec Z, H, D, J observées. Comme on voit, les valeurs $\operatorname{tg} J$ et $\sin D$ ne dépendent pas du moment magnétique du globe.

Nous allons dans la suite étudier le cas le plus simple. Nous admettons que le moment magnétique ω de la terre est constant; nous supposons ensuite que le pôle nord magnétique décrit avec une vitesse constante autour du pôle nord géographique un cercle, ayant le centre au pôle nord géographique N.

Avec ces hypothèses nous avons $\lambda = \lambda_0 + \frac{2\pi t}{T}$, où λ_0 est une constante et T — la période de la rotation du pôle nord magnétique; $\vartheta = \vartheta_0$ est une constante.

Les équations (7) et (8) donnent

$$\left.\begin{aligned}
Z &= -\frac{2\omega}{r^3}\left\{\sin\vartheta_0\sin\varphi+\cos\vartheta_0\cos\varphi\cos\left(\lambda_0+\frac{2\pi}{T}t\right)\right\}\\
H &= \frac{\omega}{r^3}\sqrt{1-\left\{\sin\vartheta_0\sin\varphi+\cos\vartheta_0\cos\varphi\cos\left(\lambda_0+\frac{2\pi}{T}t\right)\right\}^2}\\
\operatorname{tg} J &= -2\,\frac{\sin\vartheta_0\sin\varphi+\cos\vartheta_0\cos\varphi\cos\left(\lambda_0+\frac{2\pi}{T}t\right)}{\sqrt{1-\left\{\sin\vartheta_0\sin\varphi+\cos\vartheta_0\cos\varphi\cos\left(\lambda_0+\frac{2\pi}{T}t\right)\right\}^2}}\\
\sin D &= \frac{\cos\vartheta_0\sin\left(\lambda_0+\frac{2\pi}{T}t\right)}{\sqrt{1-\left\{\sin\vartheta_0\sin\varphi+\cos\vartheta_0\cos\varphi\cos\left(\lambda_0+\frac{2\pi}{T}t\right)\right\}^2}}
\end{aligned}\right\} \tag{9}$$

6. Nous allons étudier avec plus de détails les variations séculaires des élements géomagnétiques en deux points de la surface terrestre: sur l'équateur géographique et sur le pôle nord géographique.

A) Les variations séculaires de Z, H, J et D sur l'équateur géographique. Dans ce cas $\varphi = 0$ et les formules (9) prennent la forme

$$\left.\begin{aligned}
Z &= -\frac{2\omega}{r^3}\cos\vartheta_0\cos\left(\lambda_0+\frac{2\pi}{T}t\right)\\
H &= \frac{\omega}{r^3}\sqrt{1-\cos^2\vartheta_0\cos^2\left(\lambda_0+\frac{2\pi}{T}t\right)}\\
\operatorname{tg} J &= -2\,\frac{\cos\vartheta_0\cos\left(\lambda_0+\frac{2\pi}{T}t\right)}{\sqrt{1-\cos^2\vartheta_0\cos^2\left(\lambda_0+\frac{2\pi}{T}t\right)}}\\
\sin D &= \frac{\cos\vartheta_0\sin\left(\lambda_0+\frac{2\pi}{T}t\right)}{\sqrt{1-\cos^2\vartheta_0\cos^2\left(\lambda_0+\frac{2\pi}{T}t\right)}}
\end{aligned}\right\} \tag{10}$$

Comme on peut voir, Z a la période T, tandis que H a la période $\frac{T}{2}$. On peut démontrer cette propriété de H directement en substituant dans la formule (10) [exprimant H] au lieu de t la valeur $t+\frac{T}{2}$. Nous avons

$$\sqrt{1-\cos^2\vartheta_0\cos^2\left[\lambda_0+\frac{2\pi}{T}\left(t+\frac{T}{2}\right)\right]}=\sqrt{1-\cos^2\vartheta_0\cos^2\left(\lambda_0+\frac{2\pi}{T}t\right)}$$

ce qui démontre la proposition énoncée.

B) Les variations séculaires de Z, H, J et D au pôle géographique. Au pôle nord géographique on a $\varphi=\frac{\pi}{2}$. Nous avons par conséquent

$$Z=-\frac{2\omega}{r^3}\sin\vartheta_0;\ H=\frac{\omega}{r^3}\cos\vartheta_0;\ \operatorname{tg}J=-2\operatorname{tg}\vartheta_0;\ \sin D=\sin\left(\lambda_0+\frac{2\pi}{T}t\right). \quad (11)$$

Au pôle nord géographique Z, H, J sont constantes. On peut aisément voir que si l'on mesure l'inclinaison J au pôle nord géographique, on trouve la latitude ϑ_0 du pôle nord magnétique.

Le résultat reçu pour la déclinaison D [formule (11)] devient très clair, si nous plaçons d'abord le point d'observation A à une distance finie du pôle nord géographique N (Fig. 4) et puis nous faisons rapprocher indéfiniment le point A au pôle N. D'après notre hypothèse le pôle nord magnétique n tourne autour du pôle nord géographique N en suivant la circonférence CB.

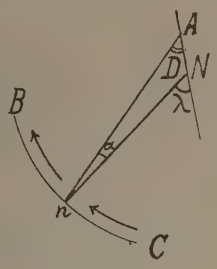

Fig. 4.

Nous avons $\lambda=\alpha+D$ ou $D=\lambda-\alpha$, d'où $D=\lambda_0+\frac{2\pi}{T}t-\alpha$. Si la distance AN décroît sans limite, l'angle α devient infiniment petit et pour la limite on a

$$D=\lambda_0+\frac{2\pi}{T}t.$$

Nous trouvons directement le résultat, que nous avons obtenu par une voie indirecte.

La comparaison des formules théoriques des variations séculaires et des résultats d'observations sera exposée dans l'article suivant.

Institut de Géophysique Théorique.
Académie des Sciences de l'URSS.

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PHYSIQUE DU GLOBE

SUR L'INTERSECTION DES ISOLIGNES

Par P. P. LASAREFF (P. P. LAZAREV), de l'Académie

§ 1. Les positions mutuelles des isolignes et leurs intersections présentent un intérêt pour la construction des cartes de magnétisme terrestre et des anomalies de la gravité. Le but du présent travail consiste en étude de cette question au point de vue de la théorie du potentiel.

§ 2. Soit F la valeur de la force au point A, placé près de la surface du globe, et X, Y, Z les composantes rectangulaires de F. X est dirigée vers le nord, Y vers l'est et Z représente la composante verticale. Nous admettons, que la surface terrestre est plane et que X, Y, Z sont les fonctions des coordonnées x, y, z du point A et dépendent du potentiel V.

Nous pouvons écrire

$$X = -\frac{\partial V}{\partial x}; \quad Y = -\frac{\partial V}{\partial y}; \quad Z = -\frac{\partial V}{\partial z}.$$

Si K_1, K_2, K_3 sont les constantes, les équations

$$X = -\frac{\partial V}{\partial x} = K_1; \quad Y = -\frac{\partial V}{\partial y} = K_2; \quad Z = -\frac{\partial V}{\partial z} = K_3 \tag{1}$$

représentent les surfaces sur lesquelles les composantes correspondantes restent constantes. Ces surfaces nous voulons nommer isosurfaces.

En coupant les isosurfaces (1) par les plans horizontaux, parallèles à la surface de la terre, on obtient les lignes d'intersection le long desquelles les composantes X et Y sont constantes. Ces lignes portent le nom des isolignes.

Les équations (1) permettent de recevoir les expressions suivantes:

$$\frac{\partial^2 V}{\partial x^2} + \frac{\partial^2 V}{\partial x \partial y}\left(\frac{\partial y}{\partial x}\right)_X = 0; \quad \frac{\partial^2 V}{\partial x \partial y} + \frac{\partial^2 V}{\partial y^2}\left(\frac{dy}{dx}\right)_Y = 0 \tag{2}$$

$\left(\frac{dy}{dx}\right)_X$ est le coefficient angulaire a_X de la tangente à une isoligne X au point A; $\left(\frac{dy}{dx}\right)_Y$ est le coefficient angulaire a_Y appartenant à une isoligne Y au même point A. Nous avons

$$a_X = \left(\frac{dy}{dx}\right)_X = -\frac{\frac{\partial^2 V}{\partial x^2}}{\frac{\partial^2 V}{\partial x \partial y}}; \quad a_Y = \left(\frac{dy}{dx}\right)_Y = -\frac{\frac{\partial^2 V}{\partial x \partial y}}{\frac{\partial^2 V}{\partial y^2}}. \tag{3}$$

L'angle entre deux tangentes à deux courbes X et Y au point A s'exprime ainsi

$$\operatorname{tg}\varphi = \frac{a_Y - a_X}{1 + a_Y a_X}. \tag{4}$$

En remplaçant les valeurs a_X et a_Y par leurs expressions (3) nous avons

$$\operatorname{tg}\varphi = \frac{\dfrac{\partial^2 V}{\partial x^2}\cdot\dfrac{\partial^2 V}{\partial y^2} - \left(\dfrac{\partial^2 V}{\partial x \partial y}\right)^2}{\dfrac{\partial^2 V}{\partial x \partial y}\left(\dfrac{\partial^2 V}{\partial x^2} + \dfrac{\partial^2 V}{\partial y^2}\right)}. \tag{5}$$

§ 3. Lorsque les isolignes X et Y sont orthogonales, les coefficients a_X et a_Y des tangentes aux courbes X et Y au point A remplissent la condition

$$a_X a_Y = -1 \quad \text{ou} \quad a_X + \frac{1}{a_Y} = 0.$$

En substituant les valeurs de a_X et a_Y prises des équations (3) on obtient

$$\frac{\dfrac{\partial^2 V}{\partial x^2} + \dfrac{\partial^2 V}{\partial y^2}}{\dfrac{\partial^2 V}{\partial x \partial y}} = 0. \tag{6}$$

La valeur $\dfrac{\partial^2 V}{\partial x \partial y}$ ne peut pas être infinie. Pour remplir la condition d'orthogonalité (6) il faut donc avoir

$$\frac{\partial^2 V}{\partial x^2} + \frac{\partial^2 V}{\partial y^2} = 0. \tag{7}$$

D'après la théorie du potentiel dans le champ qui ne contient pas des masses, les dérivées partielles du potentiel doivent remplir l'équation de Laplace

$$\frac{\partial^2 V}{\partial x^2} + \frac{\partial^2 V}{\partial y^2} + \frac{\partial^2 V}{\partial z^2} = \Delta V = 0.$$

Cette équation combinée avec l'équation (7) montre que, si les isolignes X et Y se coupent sous un angle droit, la valeur $\dfrac{\partial^2 V}{\partial z^2}$ doit être égale à zéro. Cette relation est équivalente à l'expression

$$-\frac{\partial V}{\partial z} = Z = f(x, y).$$

L'orthogonalité des deux systèmes des isolignes X et Y existe, lorsque Z est constante sur la ligne z, située au-dessus de la surface du globe.

Un cas spécial correspond à la condition $Z = 0$. Dans ce cas la force est entièrement placée dans le plan xy des coordonnées, et les isolignes des composantes X et Y sont toujours orthogonales. L'orthogonalité s'observe également lorsque

$$\frac{\partial^2 V}{\partial x \partial y} = 0 \text{ [voir formule (5)]}.$$

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CHEMISTRY

INFLUENCE OF RADIATIONS DURING THE CORROSION OF METALS ON THE DESTRUCTION OF PERIODIC PRECIPITATES OF $Ag_2Cr_2O_7$ IN GELATINE

By P. F. MIKHALEV, A. A. ULYANOV and F. M. SHEMIAKIN

(Communicated by Wl. A. Kistiakowsky, Member of the Academy, 23. VII. 1939)

The influence of various kinds of radiations on periodic reactions is very interesting, and all the more so, since this problem has been studied little up to the present time. The influence of visible light on periodic reaction has been studied by many authors (Davis, Fisher, Michalev and Shemiakin, Dogadkin, Kewster, Ishemura etc.)(1). Stempell(2) observed the nitrogen reaction to destroy Liesegang's layers (Stempell's effect).

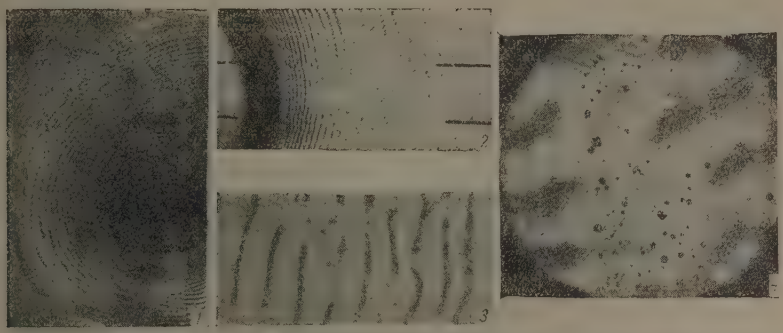
In 1936 A. A. Ulyanov(3) discovered a new kind of radiation during the corrosion of various metals: aluminium, magnesium, zinc, cadmium, lead, etc. Since these rays in all probability are in the short ultraviolet region, just as mitogenetic radiations, it was interesting to find out whether these rays have any effect upon Liesegang's rings, on the one hand, and on the other, if so, to prove the existence of these rays once more in this way.

To this purpose we chose the classic example of the formation of Liesegang's rings by silver bichromate in gelatine. Concentration of the external component silver nitrate was saturated/6. The concentration of the internal component ($K_2Cr_2O_7$) and gelatine corresponds to point No. 2 of Gibbs' triangle, i. e. 2 ml saturated potassium bichromate+3 ml water+5 ml table gelatine 10%. Prepared in this manner, this mixture was poured out in as thin a layer as possible onto a glass plate. After the gel had cooled, drops of the solution of the external component silver nitrate were applied to its surface. Glass rods 1.5 to 2 mm thick were placed at opposite ends of the plate; the ends of the preliminarily activized metallic plates which served as sources of radiations during their corrosion, rested upon these glass rods.

Aluminium and magnesium were chiefly used in the experiment, since according to A. A. Ulyanov's experiments they give the most intensive radiations. The plates of metal, freshly cleaned with emery, were activized by smearing them once with a 0.001 *n* solution of mercuric chloride(2). Made ready in this manner the preparations were placed in a humid chamber and shut off from daylight.

The irradiation lasted about 15 hours. The results obtained show that radiations of metals during corrosion destroy the rings in the region immediately under the radiating metal plate (this can be seen clearly from the annexed photograph).

A large number of experiments showed that this phenomenon of the destruction of Liesegang's rings by the corrosion radiations of metals can be brought about quite easily. In the photographs the position of the metallic radiating plates is depicted by dark lines. The most typical cases of the action of corrosion radiations on the decomposition of Liesegang's rings are photographed in Figs. 1—4. Fig. 1 shows a part of the preparation in natural size, on which the regions of destruction of Liesegang's rings can clearly be seen



to correspond to the orientation of the radiating metallic plate. It is necessary to note that the pronounced destruction of the rings immediately under the edges of the metallic plate confirms the presence of more intensive radiations from the active spots of the plate (the edges). In Fig. 2 part of the same preparation is shown magnified four times. Fig. 3 shows separate layers magnified 15 times. Fig. 4 gives a photomicrograph of a part of the same preparation magnified 52 times. The last two photographs show clearly that in the regions where the rings are ruptured, large globulites of a precipitate of silver bichromate are formed; these globulites differ pronouncedly in their degree of dispersion from the precipitations of normal layers.

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CHEMISTRY

MECHANISM UNDERLYING THE ACCELERATION AND RETARDATION OF THE AUTOXIDATION OF HYDROCARBONS. I. INFLUENCE OF ACCELERATORS AND INHIBITORS OF AUTOXIDATION ON THE DECOMPOSITION OF ORGANIC PEROXIDES

By K. I. IVANOV, V. K. SAVINOVA and E. G. MIKHAILOVA

(Communicated by S. S. Nametkin, Member of the Academy, 22. VII. 1939)

Considerable experimental material on the catalytic acceleration and retardation of the autoxidation of hydrocarbons which has now accumulated deals with the influence of positive and negative catalysts on the autoxidation processes and on combustion in general. But in order to get an adequate knowledge about the mechanism of this phenomenon, it seems more interesting to study the action of catalysts on the intermediate reactions of the oxidation process. According to modern ideas, the autoxidation phenomenon, in agreement with the Bach-Engler theory, begins with the formation of organic peroxides, through the addition of molecular oxygen to the hydrocarbon molecule. Recently these ideas have been supported by the isolation of peroxides having a definite structure from the products of the autoxidation of saturated hydrocarbons. These peroxides have been isolated in such quantities as to leave no doubt about their formation, being not a casual side reaction, but a normal intermediate stage of the oxidation and combustion of hydrocarbons and of their mixtures. Thus, the main product obtained by Hock and Susemihl (1) from the oxidation of tetralin in the liquid phase by air at 70°C, was a hydroperoxide having the structure

— OO — H.

In 1935—36 in the Laboratory for Combustibles and Oils of the All-Union Heat Institute(2) from the products of the vapour phase oxidation of cyclohexane as well as normal and iso-heptanes with air at 250—330° C there were isolated up to 30% (from the weight of the hydrocarbon having entered into reaction) of peroxide compounds, among them dialkyl peroxides and a series of hydroxymethyl peroxides. Among the latter in the products of the oxidation of isomeric heptanes the following crystalline peroxides have been found: $HO \cdot CH_2 - OO - CH_2OH$ (dihydroxymethyl peroxide) and $(C_2H_3O) - OO - CH_2OH$, having probably the structure $CH_2 - CH - OO -$

$\quad\quad\quad

compound, and under conditions of oxidation at a lower temperature, a polyatomic peroxide $C_7H_{14}O_7$, to which there were reasons to ascribe the structure $(H-OO)_2 \cdot C_6H_9-OO-CH_2OH$ [3].

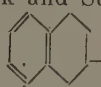
[Results confirming the preliminary formation of peroxides during the oxidation of saturated hydrocarbons and giving as well an idea about this structure have been obtained also by Mondin-Monval and Quanquin[4], Pease and Munro[5], Medvedev and collaborators[6], Harris and Egerton[7]].

Thus, it has been shown that in the course of oxidation of cyclic hydrocarbons, particularly of naphthenes, it is possible under certain conditions to observe the formation of peroxides which are the primary (relatively stable) products of the addition of molecular oxygen to the initial hydrocarbon molecule. In this case the addition is not accompanied by simultaneous splitting of the initial hydrocarbon molecule, as in the case of saturated hydrocarbons with an open chain of C-atoms which have a less stable carbon skeleton.

These data, supporting the main points of the peroxide theory, show at the same time that the predominating rôle in the mechanism of the autoxidation of hydrocarbons containing no active double and triple bonds, belong apparently to hydroperoxides and hydroxyalkyl (in particular to hydroxymethyl) peroxides.

Under the conditions of a further development of the oxidation process, these peroxides undergo a thermal decomposition. This is the second and very important intermediate stage of this process. Its significance is particularly considerable in the high-temperature oxidation (combustion) of hydrocarbons. Thanks to the evolution of a considerable amount of energy it determines the entering into the reaction of new portions of the initial hydrocarbon with the formation of new amounts of peroxides. This stage, according to the chain mechanism, is a factor in the ramification of the reaction chain.

Therefore, in order to judge about the mechanism of the action of accelerators and inhibitors on the autoxidation of hydrocarbons it seemed important to determine which of these two main intermediate stages of the process is influenced by them. Accordingly, it was necessary to observe their action on the thermal decomposition of the peroxides forming during the autoxidation.

The initial products of the investigation we undertook in that direction were the hydroperoxide of tetrahydronaphthyl (m. p. 55.8—56.0°C) prepared according to Hock and Susenmihl, and the hydroxymethyl peroxide of tetrahydronaphthyl  (m. p. 46. 8—47.3°C) obtained for the first time in our laboratory[8].

At first we studied the kinetics of the decomposition of these peroxides in solutions of tetralin and decalin, containing 4 molecular per cent of the peroxide (about 5% by weight) at temperatures from 70 to 150°C. At definite time interval samples were taken, using a special procedure, from solutions that were kept at a constant temperature in quartz vessels in a CO_2 atmosphere. In these samples the content of the peroxide that had not decomposed was determined by means of iodometric analysis in a series of experiments with the hydroperoxide, and by means of stannometric analysis in experiments with the hydroxymethyl peroxide*.

* As it was shown by the determination of the active oxygen in pure peroxides, the iodometric method in the case of $C_{10}H_{11}-OO-CH_2OH$ gives somewhat higher results,

It has been found that:

1. The velocity of the thermal decomposition of the hydroxymethyl peroxide is higher than that of the hydroperoxide at the same temperatures.

2. The velocity of the decomposition of the peroxides in both of the solutions investigated is not quite the same. In tetralin, for instance, the decomposition of the hydroperoxide proceeds somewhat more rapidly than in decalin.

3. The calculation of the constants of the decomposition velocities of both peroxides according to equations of the zero, first and second order gives better results for the equation of the first order, this being particularly clear for the hydroperoxide. But even in that case the constant is not always quite the same, this being characteristic for many-steps reactions.

4. The dependence existing between the constants of the velocities of the thermal decomposition of $C_{10}H_{11}—OO—H$ and $C_{10}H_{11}—OO—CH_2OH$ in tetralin (calculated according to the equation of the first order) and the temperature agrees with the Arrhenius equation (Fig. 1). The calculation of the energy of activation gave the following results:

for	$C_{10}H_{11}—OO—H$	24000 cal. per mol.
for	$C_{10}H_{11}—OO—CH_2OH$	26500 cal. per mol.

Hence the constant of the velocity of decomposition in tetralin of $C_{10}H_{11}—OO—H$ can be calculated from the equation:

$$k = 0.93 \cdot 10^9 \cdot e^{-\frac{24000}{RT}} \qquad \text{(time in seconds)}$$

and the velocity constant for $C_{10}H_{11}—OO—CH_2OH$ from the equation:

$$k = 0.62 \cdot 10^{12} \cdot e^{-\frac{26500}{RT}} \qquad \text{(time in seconds).}$$

In order to determine the influence of catalysts on the thermal decomposition of these peroxides, the velocity of the decomposition of $C_{10}H_{11}—OO—H$ at 140° and of $C_{10}H_{11}—OO—CH_2OH$ at 100° C has been investigated in tetralin and decalin in the presence of compounds known to be positive and negative catalysts of the autoxidation of hydrocarbons and as antiknocks and promoters of knocking in the combustion of hydrocarbon fuel in the internal engines.

The procedure for the carrying out of the experiments and for the taking of samples was the one used in the study of the decomposition of pure peroxides. The catalysts dissolved in the solvents tested (usually 1 mole per 50 moles of the peroxide) were added to the solutions of peroxides in the same solvents at the moment when the latter attained the necessary temperature. At the same time control experiments on the decomposition of the pure peroxide (without catalysts) in the same solvent were carried out. In a separate test-tube the solution of the investigated catalyst in the tested solvent (in the absence of peroxides) was kept at the temperature of the experiment. The content of active oxygen in this solution was determined periodically in order to make corrections for the capacity of the catalyst itself to react with HI or with $SnCl_2$ and its influence upon the oxidation of the solvent by the oxygen of air during the experiment.

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particularly at low concentrations of the peroxide. This cannot be obviated by the variation of working conditions, in particular by the variation of the time of contact with the HI solution. By means of a solution of $SnCl_2$ [under conditions of the procedure of Hock and Schrader (9), somewhat modified by us] the active oxygen can be determined with satisfactory accuracy (from 96 to 100% from the theoretical amount) in the pure hydromethyl peroxide as well as in its solutions in tetralin and decalin.

In Figs. 2 and 3 curves are shown that illustrate the decomposition of peroxides in the presence of some of the investigated catalysts plotted against time. A comparison of the influence of the investigated catalysts on the velocity of the decomposition of both peroxides may be found in diagrams 4, 5 and 6 where the remaining concentrations of the peroxides are given after a 1½ hour decomposition: a) in a solution containing no additions (horizontal line), b) in solutions where catalysts are present.

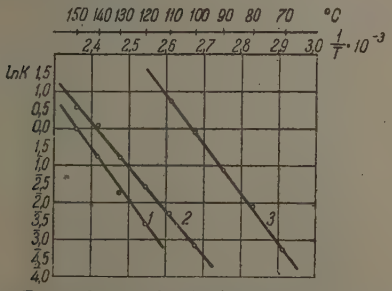

Fig. 1. Dependence of velocity constants of the decomposition of peroxides $C_{10}H_{11}-OO-H$ (*1*—in decalin, *2*—in tetralin) and $C_{10}H_{11}-OO-CH_2OH$ (*3*—in tetralin) upon the temperature expressed in co-ordinates ln k and $\frac{1}{T}$.

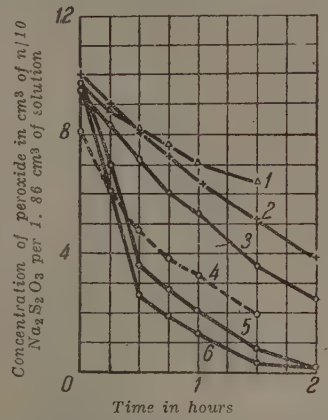

Fig. 2. Decomposition of $C_{10}H_{11}-OO-H$ in decalin at 140°C. *1*—α-naphthol (1:50); *2*—without additions; *3*—+octylnitrite (1:50); *4*—+ Fe naphthenate (1:270); *5*—+ Pb $(C_2H_5)_4$ (1:50); *6*—+ ethyl solution.

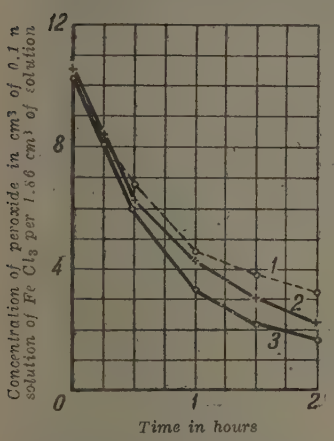

Fig. 3. Decomposition of $C_{10}H_{11}-OO-CH_2OH$ in tetralin at 100°C. *1*—+lead tetraethyl (1:50); *2*—without additions; *3*—+ octylnitrite (1:50).

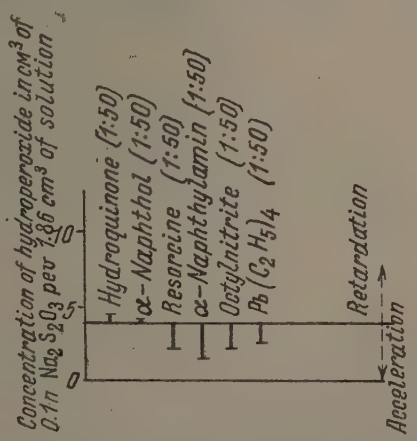

Fig. 4. Comparison of the influence of negative catalysts of the autoxidation of hydrocarbons as well as of antiknocks and promoters of knocking upon the decomposition of the hydroperoxide $C_{10}H_{11}-OO-H$ in a tetralin solution at 140°C (during 1.5 hours).

The data obtained show that:

1. The group of substances known to be negative catalysts of autoxidation in the liquid phase (hydroquinone, resorcinol, α-naphthol, phenyl-*p*-aminophenol, α-naphthylamine, phenyl-β-naphthylamine) fails to produce any definitely orientated action on the decomposition of the peroxides investigated.

In many cases these substances inhibit the decomposition, but usually only in a small degree. In some cases they have no influence at all, and at

others they accelerate the decomposition. Their action is not the same upon different peroxides and even upon the same peroxide but in different solutions. The predominating tendency of these inhibitors of autoxidation towards the hydroperoxide $C_{10}H_{11}$—OO—H is the inhibition of its decomposition, but in the majority of cases this inhibition is insignificant. This action takes place in a decalin solution. The velocity of the decomposition of the hydroperoxide in a tetralin solution is not influenced by hydroquinone and α-naphthol. Resorcinol and α-naphthylamine act as accelerators. The inhibitors that

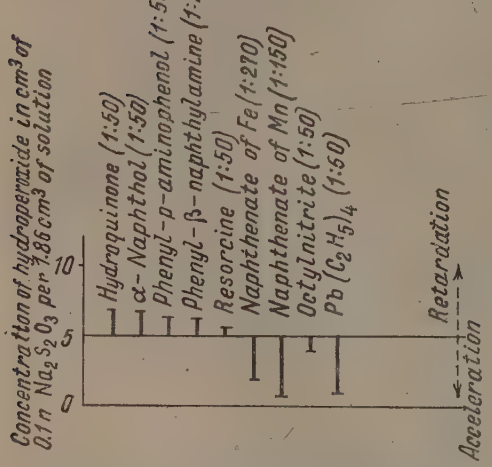

Fig. 5. Comparison of the influence of positive and negative catalysts of the autoxidation of hydrocarbons as well as of antiknocks and promoters of knocking upon the decomposition of the hydroperoxide $C_{10}H_{11}$—OO—H in a decalin solution at 140°C (during 1.5 hours).

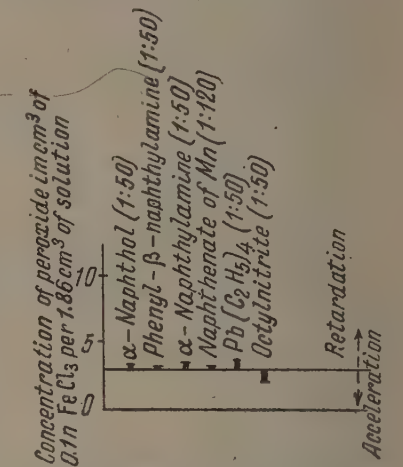

Fig. 6. Comparison of the influence of positive and negative catalysts of the autoxidation of hydrocarbons as well as of antiknocks and promoters of knocking upon the decomposition of the hydroxymethyl peroxide $C_{10}H_{11}$—OO—CH_2OH in a tetralin solution at 100°C (during 1.5 hours).

were tested have practically no influence upon the decomposition of the hydroxymethyl peroxide $C_{10}H_{11}$—OO—CH_2OH.

2. Positive catalysts of the autoxidation of hydrocarbons (Fe and Mn naphthenates) accelerate the decomposition of the hydroperoxide, but have no influence upon the decomposition of the hydroxymethyl peroxide.

3. Lead tetra-ethyl and the solution that contains it accelerate rather energetically the decomposition of the hydroperoxide and inhibit somewhat the decomposition of the hydroxymethyl peroxide. Octylnitrite has an accelerating action on the decomposition of both peroxides investigated.

4. The results obtained by us, in the part dealing with the influence of antioxidants (hydroquinone, α-naphthol and α-naphthylamine) upon the decomposition of tetrahydronaphthyl hydroperoxide disagree sharply with the data published recently by Yamada[10].

This author, on the base of his experiments on the decomposition of the same hydroperoxide in the presence of the same substances, came to the conclusion that all of the latter, at temperatures and concentrations near to those investigated by us, are strongly inhibiting the decomposition of the

hydroperoxide in tetralin. On the base of these facts this author made the conclusion that the mechanism of their action in the autoxidation of hydrocarbons consists just in their inhibiting influence upon the decomposition of peroxides occurring in the oxidation process.

This erroneous conclusion, as the results of our own experiments have shown, depended upon the conditions under which the work of Yamada was carried out. These conditions cannot be considered as garanteeing the reliability of the experimental results obtained by him. In particular, he used as the initial material for the study of the decomposition of the peroxide not solutions of the absolutely pure crystalline hydroperoxide (m. p. 55.8—56. 0°C) in pure tetralin as it was done by us, but non-purified crude solutions of the hydroperoxide obtained as a result of a prolonged oxidation of tetralin with air. The latter, under such conditions, are always contaminated with resinous substances and with other products of the deep oxidation of tetralin, as well as with impurities found in tetralin. All this, without doubt, has told on the results obtained by Yamada.

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CHEMISTRY

MECHANISM UNDERLYING THE ACCELERATION AND RETARDATION OF THE AUTOXIDATION OF HYDROCARBONS. II. INFLUENCE OF ACCELERATORS AND INHIBITORS OF AUTOXIDATION ON THE FORMATION OF ORGANIC PEROXIDES

By K. I. IVANOV, V. K. SAVINOVA and E. G. MIKHAILOVA

(Communicated by S. S. Nametkin, Member of the Academy, 22. VII. 1939)

In a previous paper (see Communication I) we have established the absence of a definite direction and of any considerable activity in the action of inhibitors of the autoxidation of hydrocarbons upon the reaction of the thermal decomposition of hydro- and hydroxymethyl-peroxides of tetrahydronaphthyl in organic solvents. This as well as other facts such as, for instance, a similar influence on this process of antiknocks as well as of promoters of knocking that was observed in many cases, lead us to the idea of studying the action of all these substances upon an earlier stage of the oxidation process, viz. upon the reaction of the initial formation of peroxides during the autoxidation process.

To that effect the influence of these substances, as well as of Fe and Mn naphthenates, upon the velocity of the formation of that same hydroperoxide of tetrahydronaphthyl during the low temperature oxidation of tetralin with molecular oxygen was studied. This latter reaction [its kinetics and mechanism have been recently investigated by S. Medvedev and A. Podyapolskaya([1])] proceeds at low temperatures only to the stage of the formation of the hydroperoxide $C_{10}H_{11}—OO—H$([2]).

Therefore, under these conditions one can expect to trace the influence of autoxidation catalysts just on the reaction of the initial formation of the peroxide which would not be considerably complicated by its simultaneous decomposition as it is usually the case.

The oxidation of tetralin was carried out in an apparatus consisting of a cylindrical vessel (diameter 20 mm, length 325 mm) fitted with a ground-glass effective reflex condenser. In the lower part of the vessel an inlet tube was sealed in for the oxidizing gas. On the base of preliminary experiments on the oxidation of pure tetralin in this apparatus we developed the following conditions of the process that warranted a sufficient velocity of the formation of the hydroperoxide and a satisfactory squaring of the results. These conditions were: the bubbling of pure dry oxygen through 20 g of the hydrocarbon maintained at 70° C at a rate of 250 cm³ per minute.

The apparatus containing each 20 g of a solution of the catalysts that were tested in pure distilled tetralin ($d_4^{15}=0.972$, $n_D^{19}=1.5418$) were placed in a water bath where the temperature was kept at $70\pm0.5°$ C, and a current of

dry O_2 was passed through the solution. At definite time intervals samples of 0.96 cm³ were taken and the content of the hydroperoxide was determined using the iodometric method. Each series of experiments with catalysts was followed, for comparison's sake, by the oxidation of a sample of the tetralin used (in the absence of catalysts).

The curves showing the dependence of the velocity of formation of the hydroperoxide on the presence of some of the catalysts investigated are given in Fig. 1. The duration of the oxidation of tetralin, expressed in hours, is plotted on the abscissa axis, and the increase in the concentration of the hydroperoxide expressed in cm³ of a 0.1 *n* solution of thiosulphate per 1 cm³ (more accurately per 0.96 cm³) of the solution is plotted on the ordinate axis.

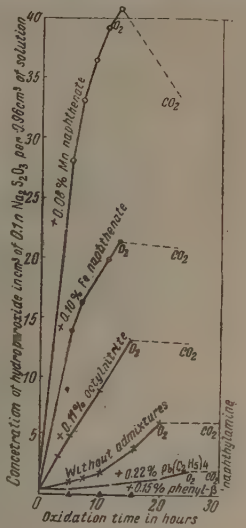

Fig. 1. Influence of some positive and negative catalysts of the autoxidation of hydrocarbons as well as of antiknocks and of promoters of knocking upon the velocity of the formation of the hydroperoxide $C_{10}H_{11}$—OO—H during the oxidation of tetralin (t=70°C).

In order to check whether simultaneously with the formation of the hydroperoxide at the temperature of the experiment (70°C) its partial decomposition takes place as well, the heating at the end of the oxidation was continued for some time, the oxygen being substituted by carbon dioxide. After some time the active oxygen in this reaction mixture was determined (a dotted line prolongation of the curves in Fig. 1) which showed that in the oxidation of tetralin, in the absence of added substances as well as in the presence of inhibitors of the formation of the hydroperoxide, there is no decomposition of the latter. The accelerators of its formation: octylnitrite and particularly Fe and Mn naphthenates call forth a certain decomposition of the peroxide formed. Therefore, their absolute influence on the velocity of the formation of the hydroperoxide would be in those cases even more effective than it may be seen from the curves.

A graphical comparison of the influence of all the substances tested on the formation of $C_{10}H_{11}$—OO—H during the autoxidation of tetralin is given in Fig. 2. This diagram is plotted according to the same principle as the comparative diagrams given in Communication I, that summarize the data concerning the action of all the catalysts investigated on the decomposition of peroxides.

The results obtained show that:

1. The inhibitors of liquid phase autoxidation of hydrocarbons (hydroquinone, resorcinol, α-naphthol, phenyl-*p*-aminophenol, phenyl-β-naphthylamine and α-naphthylamine) inhibit very strongly the initial formation of the hydroperoxide of tetrahydronaphthyl formed during the oxidation of tetralin, stopping practically completely its appearance within the range of oxidation time investigated by us.

2. Naphthenates of Fe and Mn, that are considered to be positive catalysts of the oxidation of hydrocarbons, proved to be extremely active accelerators of the formatior of the hydroperoxide. In the presence of 0.08% of Mn naphthenate, for instance, a fourteen times larger amount of the peroxide was formed compared to the oxidation of pure tetralin during the same time interval (13 hours).

3. Lead tetra-ethyl inhibits considerably the formation of $C_{10}H_{11}$—OO—H on the oxidation of tetralin, while octylnitrite accelerates it considerably.

4. The comparison of these data with the results of the study of the action of the same substances upon the decomposition of the hydroperoxide $C_{10}H_{11}$—OO—H as well as of the hydroxymethyl peroxide $C_{10}H_{11}$—OO—CH_2OH in organic solvents, given in Communication I, allows us to state the following:

a) there exists no parallelism between the direction (or strength) of the action of the catalysts investigated on the oxidation of hydrocarbons, as

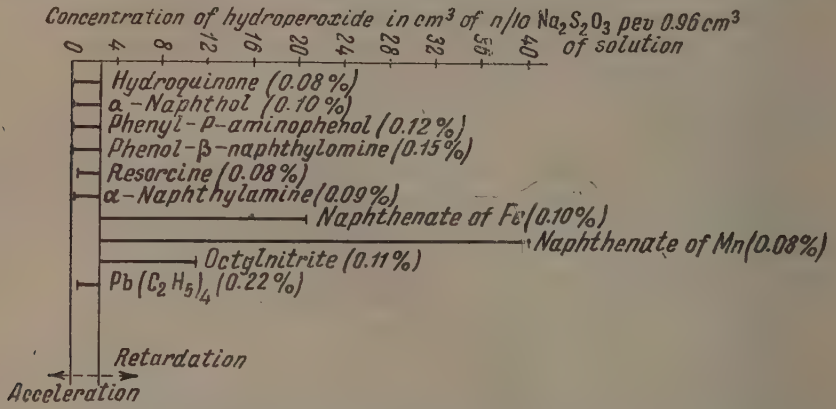

Fig. 2. Comparison of influence of positive and negative catalysts of the oxidation o hydrocarbons as well as of antiknocks and of promoters of knocking upon the formation of the hydroperoxide $C_{10}H_{11}$—OO—H during the oxidation of tetralin with oxygen at 70°C (duration of the experiment about 13 hours).

reported in literature, and their influence upon the decomposition velocity of the peroxides investigated (including those that are formed during autoxidation);

b) there exists a complete parallelism between the direction of the action of these substances on the total velocity of the oxidation process and the character of their influence on the velocity of the reaction of the formation of peroxides during autoxidation.

The conclusion to be drawn is that the mechanism of the action of accelerators and inhibitors of the autoxidation of hydrocarbons upon the oxidation process, at least when the latter proceeds at comparatively low temperatures in the liquid phase, is not in their main influence being directed upon the decomposition reaction of the peroxides that are formed as intermediates, in agreement with wide spread ideas, particularly developed by Moureu, Yamada and others, but it is the reaction of the initial formation of these peroxides in the autoxidation processes which is largely controlled by them.

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BIOCHEMISTRY

BIOCHEMICAL PECULIARITIES OF VARIOUS REGIONS OF AXOLOTL'S BODY CONNECTED WITH FORM-BUILDING UNDER THE ACTION OF ORGANIZERS

By A. N. ADOVA and A. M. FELDT

(Communicated by N. D. Zelinsky, Member of the Academy, 8. VII. 1939)

The present work in its first stage attempts to settle the question as to whether there exists a difference in proteolytic faculty of the skin, taken from various morphogenetic regions of the axolotl's body, and whether this faculty is affected by the presence of tissues—«organizers». The presence of a larger number of proteases or their higher activity must exert the influence on a simple autolysis, and also in experiments with participation of «organizers».

With this object in view we have performed a series of experiments on autolysis of the skin, taken from the extremities and from the back—two regions which differ the most in their capacity to give growth to adventitious formations. The intensity of the proteolytic processes was determined from the increase of aminonitrogen and COOH groups for the periods of 20 and 70 hours. Part of the experiments was conducted in the presence of cartilage and liver. The decomposition of component elements of the skin in autolysis and in the presence of preparations of cartilage and liver was studied also by the histological method (communication will follow).

The skin on the back and on the extremities shows considerable difference even if observed macroscopically. The layer of collagenic fibres on the back is several times its thickness on the extremities: 1 mm as against 0.25 mm. Determinations of moisture, ash and total nitrogen have revealed differences in relative quantities of these ingredients.

One should notice the constant difference between the fore and hind parts. Thus, the fore part of the back shows 0.520%, the hind—0. 551%; the fore extremities—0.580%, the hind ones—0.623%.

Autolysis Tests of the Skin. Aliquots of skin (without preliminary reduction to fragments) were immerged into toluene and placed into thermostat at +38°C for 70 hours. After that the toluene was drained off and the tissue was levigated with distilled water in proportion of 1 : 9. The obtained suspension was centrifuged, and aminonitrogen (micro Van Slyke) and carboxyl groups (Willsteimer) were estimated in the extract. The control estimations were made in extracts without preliminary keeping in the

Table 1

Skin Composition of Various Regions of Axolotl's Body

(% in the fresh tissue)

Region of the skin	Moisture	Dry remainder	Ash	Organic substance	Nitrogen
Skin of the back	92.0	8.0	0.52	7.46	1.09
Skin of the extremities	89.1	10.9	0.60	10.30	1.01

thermostat. The increase in aminonitrogen and acidity was calculated in mg-equiv. per 100 cm³ of extract from the difference between the control aliquots and those kept in the thermostat 70 hours. This was the procedure adopted in the first and the second series of tests. In the last series of tests (No. 3, 4 and 5) the skin aliquots were in a preliminary way levigated with citrate buffer solution in the proportion of 1 : 9, and then part of the samples were subjected to immediate analysis, while the remaining samples were placed in thermostat for 20 and 70 hours, whereupon they were analyzed.

Table 2

Increase in Acidity and in Aminonitrogen on Autolysis of Skin Taken from Various Regions of the Body of an Axolotl

(in mg-equiv. per 100 cm³ of extract)

No. of series		Skin from the back				Skin from the extremities			
		20 hours		70 hours		20 hours		70 hours	
		Acidity	Aminonitrogen	Acidity	Aminonitrogen	Acidity	Aminonitrogen	Acidity	Aminonitrogen
1	6.8	—	—	0.076	0.00	—	—	0.592	0.70
2	6.8	—	—	0.076	0.00	—	—	0.497	0.40
3	5.9	0.00	0.04	0.00	0.23	0.425	0.38	0.737	0.90
4	4.7	0.07	—	0.16	—	0.67	—	1.50	—
5	4.7	0.18	0.13	0.31	0.21	0.46	0.49	1.20	1.06

The autolysis tests make it possible to establish that proteolytic processes in the skin of extremities are considerably more intensive as compared to the skin of the back.

Proteolytic processes in mixture of skin with cartilage and liver were run in the same manner as the autolysis tests, the only difference being that the skin was mixed with cartilage or liver in proportion of 5 : 1 and was then levigated in a buffer solution with pH = 6.0. In the control test with the cartilage or liver only, the dilution was such as to make the concentration of these tissues equal to that in the main tests, i. e. 1 g of liver or cartilage per 54 cm³ of buffer.

Data characterizing the increase of acidity in tests with participation of liver are presented in Table 4.

The data on the increase of acidity in mixtures of the skin with liver have brought into light the very different course of biochemical processes in dif-

Table 3

Acidity and Aminonitrogen Content Increasing in a Mixture of Skin and Cartilage (in mg-equiv. per 100 cm³ of extract)

No. of series	Skin from the back+cartilage		Skin from the extremities+cartilage		Remarks
	Acidity	Amino-nitrogen	Acidity	Amino-nitrogen	
6	1.0	—	2.21	—	Tests No. 6 and 7 were performed with addition of cystein, amount of cartilage indefinite
7	0.22	0.49	1.77	1.74	
8	0.07	0.00	0.59	0.46	Tests No. 8, 9 and 10 were performed without addition of cystein, amount of cartilage definite—1 part to 5 parts of skin. Tests with cartilage only did not produce any increase in acidity and aminonitrogen
9	0.00	0.00	0.78	0.53	
10	0.00	0.00	1.17	0.54	

Table 4

Increase of Acidity in a Mixture of Skin with Liver (in mg-equiv. per 100 cm³ of extract; pH=6.00)

No. of series	Skin from the back + + liver		Skin from the extremities + liver		Liver	
	20 hours	70 hours	20 hours	70 hours	20 hours	70 hours
11	0.63	—	0.80	—	0.13	—
12	0.18	—	0.91	—	0.13	—
13	—	0.06	—	2.59	—	0.61
14	—	0.00	—	2.13	—	0.73
15	—	0.00	—	2.80	—	0.43

ferent regions of the skin. The difference, evidently, depends upon the peculiarities of the skin, since the same liver was used throughout the experiment. If the latter was mixed with the skin from the extremities, processes leading to the increase of free COOH groups were observed, while in the mixtures with the skin from the back, even the processes of glycolysis, so characteristic of the liver, were stopped.

Nassonov made a surmise that the differences between various regions of the axolotl's skin may be accounted for by the special properties of the collagen. In order to determine the peculiarities of the collagen, it was isolated from the skin of the back and of the extremities, and its isoelectric point was established by the method of swelling. Minimum swelling was obtained at pH=4.5 in a series of tests, conducted at pH intervals equal to 0.5. The splitting up of collagen by the proteases of the tissue was determined in mixtures of protein and liver (glycerine extract) in terms of aminonitrogen. No difference in splitting behaviour of collagen whether from the back or from the extremities was observed in this test.

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GEOLOGY

SOME ISOMORPHIC SUBSTITUTIONS IN APATITE

By DUNCAN McCONNELL

(Communicated by A. E. Fersman, Member of the Academy, 9. V. 1939)

A most interesting concept of the structure of carbonate-apatites has recently been presented by N. V. Belov(1) as a modification of the hypothesis advanced by the writer in collaboration with J. W. Gruner(2,3). The paper by Belov contains numerous details, to be sure, but presents one more fundamental question, namely, the possibility of the substitution of C for Ca in the structure of apatite. This question must be considered in terms of the available experimental data.

It is regrettable that Belov did not consider the results of experimental work on the apatite group but, nevertheless, his paper presents certain valuable philosophical considerations regarding the structure. He has shown that a structure is conceivable in which decomposition of tetrahedra takes place to permit the formation of CO_3 groups (with triangular configuration), in which each CO_3 is more or less remotely associated with a F or OH ion, and that the latter occupies a fourth apex of the tetrahedron. The product attained being a distorted tetrahedron, O_3–C—(F, OH), wherein the short dash indicates an intimate association of the C and the three O ions, and the longer dash indicates a more remote association of the C and F or OH.

This hypothesis has many pleasingly commendable features on first consideration, but there may be several objectionable counterparts as this paper will attempt to point out. At any rate it does not modify the hypothesis developed by Dr. Gruner and the writer for the substitution of C for P, except that it requires that one apex of our CO_4 group must contain F or OH rather than O. This is true because no particular claim was made for any symmetry the CO_4 group might possess, except that it should be contained in the lattice in such a way as to satisfy requirements of C_{6h}^2.

On theoretical grounds Belov has accounted for the plausibility of the formula originally set down by Borneman-Starinkevitch(4), $Ca_5C_3O_9(F,OH)_4$, but no attempt was made to justify this formula in terms of the analyses of carbonate-apatites. However, the final acceptance or rejection of any hypothesis must be based upon the manner in which it fits the experimental data. This formula, were it to fit the experimental data, would be the nicer formula by far, in that it does not involve commas within the brackets and is more simple in other respects than the one which Dr. Gruner and the writer originally postulated. For this reason Borneman's formula has certain requisites which deserve special examination.

For every distorted or «decomposed» tetrahedron which contains C there must be one associated F or OH to obtain the required O_3–C—(F, OH) configuration. This is distinctly not in accord with the writer's results for dahllite[5], in which case only 0.657 ions of F+OH+Cl are available to satisfy 0.990 C ions. In francolite[2], on the other hand, 3.168 F+OH are present to satisfy only 0.765 C ions. In francolite analysed by H. C. G. Vincent[6] there are 2.013 ions of F+OH to satisfy 0.462 C ions, and in the francolite analysed by E. B. Sandell[7] 2.50 ions of F+OH are available to accommodate the 0.78 C ions. In all instances the two fluorine ions necessary to accommodate the usual fluorine positions must be deducted, and it seems essential to point out that these numbers are not greatly altered by the normalizing factor

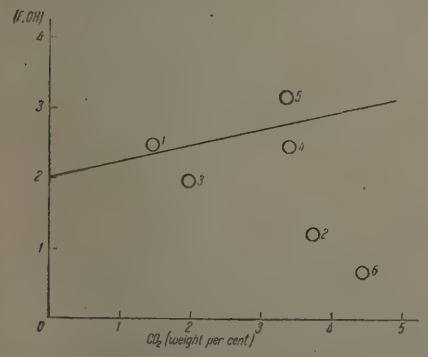

Fig. 1. Comparisons of CO_2 and (F, OH) for carbonate-apatites analysed by: 1) Y. V. Morachevsky [9], 2) E. Y. Rode [9], 3) H. C. G. Vincent [6], 4) E. B. Sandell [7], 5) R. B. Ellestad [2], 6) Ledoux and Co. [5].

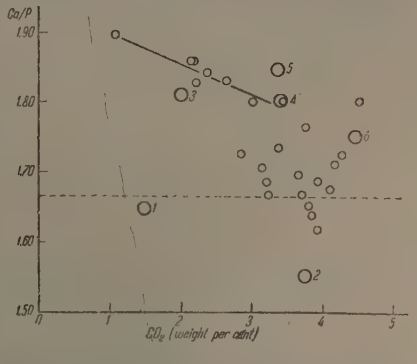

Fig. 2. Comparisons of CO_2 and Ca/P for teeth and bones and for the six mineral substances of Fig. 1.

that is used. They are relatively correct irrespective of the normalizing factor, and it is also true that the normalizing factor does not differ from unity by more than 2 or 3% in most instances. In all of these cases the amount of F+OH+Cl ions is less than the amount which the C ions would require, with one notable exception in which case it is considerably greater. For the present it will suffice to conclude that there is no simple relation between the amount of C and F+OH, although this should not be true if Borneman's formula were sufficient to account for the end-member of this sort of isomorphic substitution.

The formula previously advanced by Dr. Gruner and the writer does not require any conformity of the amount of C and (F, OH) and, indeed, the available analyses of carbonate-apatites do not show any semblance of conformity in this respect. In Fig. 1, six of the more recent and more complete analyses available in the literature have been plotted. The sum of the ionic amounts of F+OH+Cl are shown as the ordinate and the weight per cent of CO_2 as the abscissa. These are taken directly as yielded by the analyses, no normalizing factors having been applied. The line on the diagram indicates the trend of (F, OH) which is required by Borneman's formula, i. e., an increase of (F, OH) directly proportional to the increase in C ions. Nothing of the sort is exhibited by the analyses, so it must be concluded that Borneman's formula does not fit these results even in an approximate way.

Biochemists have calculated the Ca/P ratios for a considerable number of analyses of teeth and bones during the past several decades, and these values might be expected to throw some light on the matter. This seems highly pro-

bable because, if C substitutes for Ca, there should be some reflection of this fact in the Ca/P ratio. It merely remains to recognize this criterion and interpret it. Increases in the Ca/P ratio above the theoretical value, 1.667, may be interpreted in two ways: 1) the amount of P has been decreased by the occupation of P positions by C, or, 2) the amounts of Ca and P have both been decreased by the substitution of C for both of these ions, but there has been a proportionally greater amount of substitution of C for P. Decreases of Ca/P (below the theoretical value) can likewise be interpreted in two ways: 1) C has replaced Ca to an extent proportionally greater than the amount of replacement of P by C, or, 2) some other metallic ion, such as Na, has not been taken into account. This might result from the use of incomplete determinations where the presence of other metallic ions was not recognized. R. Klement([8]) gives the results of the determination of Ca, Mg, Na, K, P and CO_2 for a considerable number of carbonate-apatites of organic origin and, as these constitute the greatest number of analysed carbonate-apatites, it becomes necessary to use these results. This is done in spite of the realization that mineral analyses would be preferable, because comparatively few recent mineral analyses are available for apatites high in CO_2. The Ca/P ratio is plotted as the ordinate and CO_2 is plotted as the abscissa in Fig. 2.

Most of the Ca/P values are greater than theoretical value (indicated by a dotted line), but a few are less than the theoretical value. The smaller circles represent analyses of materials of organic origin and these have not been corrected, i. e., no allowance has been made for the fact that they contain Mg, Na and K. Because these constituents were not determined in every instance, all of the analyses are shown on the same basis. The fact that these values are uncorrected is not as serious as might be expected, however, because these substances do not show pronounced variations in Mg, Na and K. Six mineral analyses are given also and these have been fully corrected, carbon having been purposely omitted from consideration, because this constituent is presumed to be the balancing factor which restores the ratio to 1.667 in each case by its ability to enter the structure in either the Ca or P positions, as required to establish electrostatic neutrality. Rare earths, alkalies, alkali earths and ferrous iron have been assigned to Ca positions, while aluminum and ferric iron are assigned to P positions. Analyses high in SiO_2 or SO_3 were purposely omitted from consideration, as were analyses for which the corrections would be large for Na_2O, Al_2O_3, etc.

In a series of analyses by W. Wischert, reported by Klement([8]), it is noticeable that there is a decrease in the Ca/P ratio with increasing CO_2 content and these points are connected by a line in the diagram. This is precisely the opposite of what one would expect if all C were substituted for P in the lattice.

Fig. 2 also indicates that two analyses of minerals yield Ca/P values less than the theoretical as well as several substances of organic origin. These are based on analyses reported by A. K. Boldyrev([9]). This situation cannot be accounted for on the basis of Borneman's hypothesis, which will not permit a decrease of the number of Ca below 10 or an increase of P above 6. This is true provided that there are no vacancies in the Ca positions of the lattice and the density determinations recommend this conclusion as valid. However, according to the hypothesis advanced by the writer, there would not be any such restrictions placed upon the Ca/P ratio, and values both greater and less than the theoretical value are entirely consistent.

The hypothesis which the writer advocates was an outgrowth of a consideration of nearly all of the available analyses of carbonate-apatites. It was proposed on the basis of analytical determinations and not directly from theoretical considerations, but it has been examined in light of the structure of

apatite and has not been found inconsistent in so far as this structure is understood. Belov has discussed the substitution of C for Ca in terms of Pauling's polyhedra and considers several possibilities. These criteria were considered by the writer in terms of a structural model of apatite in which the relations are even more apparent than they are in Belov's diagrams. It can easily be seen that there must be occasionally some minor concentration of charge in association with those trigonal prisms where substitution of C for Ca takes place. These local concentrations of charge are unavoidable in defect structures, and the fact that they occur here is hardly more alarming than their occurrence elsewhere. As there is considerable evidence to support the portion of our original hypothesis which postulates C substituting for Ca and no direct evidence of a negative character, it may be assumed that this sort of isomorphic substitution is real. It might be well, then, to examine the formula which is required by these conditions.

The general formula, involving Ca, P, C, (F,OH) and O as the only constituents, may be written

$$Ca_6(P_{6-y}C_y)[O_{26-x}(F, OH)_x](Ca_{4-z}C_z)$$

where

$$x \leqslant 8, \quad y \leqslant 6, \quad z \leqslant 4; \qquad x \geqslant 0, \quad y \geqslant 0, \quad z \geqslant 0.$$

The upper limits of y and z are determined by symmetry considerations in the apatite structure. The upper limit of x is the value obtained when all of the P positions, but none of the Ca positions, contain C and electrostatic neutrality prevails. Stability conditions undoubtedly impose further limitations, but the extent to which these forces are operative cannot be ascertained from the comparatively few analyses available at present.

If only the special cases involving integral subscripts are considered, it is possible to write 23 variations of the general formula. A few of these are given as examples in order to demonstrate the extreme flexibility of the formula. The C and (F, OH) contents and the Ca/P ratios also indicate considerable variability.

	Ca/P	C ions	(F,OH)
$Ca_{10}F_2(P_2C_4)(O_{20}F_4)$	5	4	6
$(Ca_9C)(FO)(P_5C)O_{24}$	9/5	2	1
$(Ca_8C_2)F_2(PC_5)(O_{23}F)$	8	7	3
$(Ca_7C_3)F_2C_6O_{24}$	∞	9	2
$(Ca_6C_4)O_2C_6O_{24}$	∞	10	0

It is noteworthy that two of these examples, the second and the last, indicate substitution of O for (F, OH). This is the situation encountered in voelckerite or oxy-apatite, analyses of which frequently show CO_2. By assuming a reasonable Ca/P ratio and employing fractional subscripts it is possible to obtain another interesting example of the general formula, namely,

$$Ca_6[O_2(P_{4.8}C_{1.2})O_{24}](C_{1.6}Ca_{2.4}).$$

It is apparent that the 42 positions of the lattice have been accommodated, the F positions being occupied by O ions, and that C is distributed between the Ca and P positions. Electrostatic neutrality can be seen to obtain and there are no objectionable features from theoretical considerations. However, this need not represent the end-member of voelckerite, as constituents other than C may be present. This is furnished merely as another illustration of the general formula postulated above. It has been assumed for some of these examples that one O substitutes for one F in the lattice to produce voelckerite. The only alternate possibility would require vacancies in half of the F positions, but this is untenable from a consideration of the size of the ions

as well as their strategic position in the lattice. The existence of voelckerite is so well established at this juncture that it seems reasonable to omit any discussion of the matter.

It is concluded that the formula proposed here will account for an oxyapatite containing CO_2 as well as an apatite containing an amount of (F,OH) greater than is required to fill the F positions. It will account for Ca/P ratios both greater and less than the theoretical value regardless of the CO_2 content. It does not require any direct correlation between the Ca/P ratio and the CO_2 content. Borneman's formula, on the other hand, can account for only the Ca/P values greater than the theoretical value and only for those carbonate-apatites which contain (F, OH) in amount greater than is required for the fulfilment of the F positions. Furthermore, the latter formula indicates that a direct correlation between the amount of C and (F, OH) should exist, but there is no evidence of this relation in the analytical results available for consideration.

As no particular claims are made for the symmetry of the CO_4 group, it does not necessarily matter whether the group is considered as O_3—C—O or CO_4; the co-ordination of four oxygens by carbon results. The local concentration of charge caused by the substitution of C for Ca is not a serious objection to the hypothesis. Other more serious objections confront the only substitute hypothesis which has appeared, and these concern the manner in which it fits the experimental results.

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GEOLOGY

MINERALS OF THE YELLOW RIVER PLACERS

By A. L. ZAGIANSKY

(Communicated by A. E. Fersman, Member of the Academy, 28. VIII. 1939)

In connexion with the investigation of the Krivoy Rog region for rare and non-ferrous metals (2) the author has studied the composition of the placer of the Yellow river (Shu-District, Yellow river).

The mineralogical composition of the black sands taken chiefly from test-pits sunk in the river valley was found to be as follows: monazite, zircon, cassiterite, gold, topaz, tourmaline, garnet, rutile, ilmenite, sphene, anatase, magnetite, haematite, limonite, feldspar, hornblende, pyroxene, epidote, pyrite, quartz.

Monazite—prevalence of grains of yellow-orange colour; of rarer occurrence are grains of dirty-green colour; medium rounded.

Zircon—grains of white and pink colour, intensely rounded.

Cassiterite—black and brownish grains with semimetallic lustre, medium rounded.

Topaz—grains of white colour with intense vitreous lustre, strongly rounded.

Tourmaline—in prismatic grains, of dark-green colour, with typical striation along the face of the prism.

Rutile—observed in black sands in columnar grains.

In the black sands of the Yellow river two principal genetic mineral varieties have been observed: minerals of a metamorphic series, of iron ore-formation, and minerals of pegmatite veins. The most interesting rare minerals are: monazite, zircon, cassiterite and topaz; they are rounded and transferred from the granite areas with pegmatite and aplite veins lying higher up the Yellow river.

The most important geological reasons to investigate the Krivoy Rog region for rare elements were the following: first, data pointing to the Saksaghan rock series as being influenced by the granite magma; secondly, the present author's interpretation of a part of the surrounding magmatic rocks (largely granites) as relatively young ones compared to the Saksaghan series; and, lastly, adequate analysis of these data as brought together with the results of the theoretically valuable investigations in the field of geochemistry of hydrothermalites and pegmatites which appeared in this country during the last few years (A. E. Fersman).

It should be noted that the conceptions that prevailed formerly with regard to the genetics of the crystalline rocks of the Ukraine massif (in particular those of the Krivoy Rog) failed to give any theoretical ground for the

search of rare elements in that region. To these views should be referred the neptunic ideas of Barbot de-Marny ([1]), Kontkevich ([3]), Mikhalsky ([4]), Morazevich ([5]), Feofilaktov ([6]) etc.

The advocates of the old theories referred all the granites to very ancient (pre-Huronian), deeply eroded rocks, devoid of magmatic emanations.

The data recently obtained point to the necessity of theoretical and practical investigation of the particularities of distribution and concentration of rare elements, chiefly of those found in placers and primary pegmatite formations of the Ukrainian crystalline massif.

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GENETICS

AMPHIDIPLOIDS OF HARD WHEAT AND EINKORN PRODUCED THROUGH COLCHICINE TREATMENT

By A. R. ZHEBRAK

(Communicated by N. I. Vavilov, Member of the Academy, 13. VI. 1939)

Species of wheat notably differ in chromosome numbers and form a polyploid series with the basic number 7. Einkorn (*Triticum monococcum*) appears to be genetically the most isolated species ($2n=14$).

As shown by the studies by Bejerink, G. Vilmorin, Körnicke, Vavilov [1] and others, hybrids between *Tr. monococcum* and other species both of the 28 and the 42-chromosome series are completely sterile. Sterility of the hybrids of the first generation prevented introduction of this ancient species into the breeding process; it is of certain value, however, owing to its resistance to various diseases.

The method of restoration of fertility through doubling chromosome number in sterile hybrids as worked out by classical genetics and successfully applied to a number of other objects made it possible to solve this problem in the case of wheat. However, in spite of the theoretical clarity of the problem genetics did not dispose until recently of a method which would guarantee success in 100% of cases.

The works by Blakeslee and Avery [2] opened wide possibilities to production of polyploids in plants by means of colchicine treatment.

We worked persistently during several years on the problem of overcoming sterility in the hybrids between *Triticum Timopheevi* and other wheat species: it was just natural to make use of the method by Blakeslee and Avery inasmuch as colchicine treatment gave positive results in their work. At the present time we are already in possession of several tens of amphidiploids between *Tr. durum* and *Tr. Timopheevi* ($2n=56$) and of amphidiploids *Tr. durum* × *Tr. monococcum*. But it is only those between the hard variety *leucurum* and the einkorn which are delt with in the present paper.

In order to produce amphidiploids, hybrid seeds *Tr. durum* × *Tr. monococcum* were soaked during 4 hours in a 0.1% colchicine solution after formalin treatment. We disposed only of 4 hybrid kernels, one of which was preserved for control and the remaining three were subjected to colchicine treatment. After the treatment the seedlings were planted into wet quartz sand contained in Petri dishes. Only one plant survived from these three kernels treated in the above manner; it was very much behind in development and matured in the laboratory as late as in October. The colchicine treated plant bore only 4 ears, 3 of which were sterile and one ear containing

14 productive spikelets had 6 well filled kernels and one meagre one which proved later to be unviable.

The control hybrid *Tr. durum*×*Tr. monococcum* was well developed; it attained 147 cm in height, had 12 ears with 222 spikelets but they contained not a single kernel. Full sterility of the control plant and presence of an ear with seeds together with sterile ears on the experimental plant suggested the occurrence of amphidiploidy. The ears were well developed, some of them containing as many as 22 spikelets. Judging from the development of the hybrid plant one might suppose that there occurred heterosis in this cross and it was fully fixed by amphidiploidy. The plant obtained from the colchicine treated seed attained 128 cm in height. The two hybrid plants of the first generation *Tr. durum*×*Tr. monococcum* had non-pubescent spikelet glumae; their edges were coloured black. The coloration came from *Tr. monococcum*. Presence of kernels in all the well developed spikelets demonstrated full fertility of the amphidiploid.

Fig. 1. Seeds: *1*—*Tr. durum*; *2*—amphidiploid *Tr. durum*×*Tr. monococcum*; *3*—*Tr. monococcum*.

The length of the amphidiploid ear was equal to 6.5 cm. There were altogether 18 spikelets in the ear; 14 of them were fertile. Density of the ear was equal to 2.7. Kernels were asymmetrical and somewhat meagre (Fig. 1). The weight of the 6 kernels was equal to 1.625 g. The spikelet glumae had triangulate somewhat obtuse dent. Spikelets were arranged in a straight line. The side of the ear was broader than the face.

Cytological investigation was conducted on slides made of rootlets taken from treated hybrid, from fertile ear seed plants and from the control plant. The control plant of the first generation *Tr. durum*×*Tr. monococcum* contained, as should be expected, 21 chromosomes. The chimeral plant presented an entirely different aspect. Three rootlets of this plant were fixed; one of them proved to contain sectors with different chromosome numbers, namely, with 21 and 42 chromosomes. The remaining two rootlets contained 42 chromosomes. The amphidiploid nature of this plant which possessed an ear with seeds was therefore proved by these investigations. However, an additional cytological study of the plants obtained from the seeds of the chimeral plant was also conducted. As should be expected, they proved to possess 42 chromosomes (Fig. 2).

Fig. 2. *a*—*Tr. durum* × *monococcum* (amphihaploid) $2n=21$; *b*—amphidiploid *Tr. durum*× × *Tr. monococcum*, $2n=42$.

In the winter 1938—39 plants were grown from 2 kernels; they died. The remaining kernels were sown in the spring and now we are in a position to describe the other characters. When the amphidiploid is compared with the two parental forms, the difference in the coloration and width of the leaves becomes prominent. The leaves of the variety under experiment (*Tr. durum* v. *leucurum*) under the same conditions of cultivation are of light-green coloration and considerably broader than those of einkorn. The leaf colour in einkorn is dark-green under the same conditions of cultivation. The leaf of the amphidiploid is broader

than that in hard wheat and its colour is dark-green similar to that in einkorn (Fig. 3).

The present study has proved the experimental production of a 42-chromosome wheat and thus contributed to the knowledge of the origin of the whole 42-chromosome series.

As early as in 1938 Komarov wrote that there is no plant as yet which could be presumptively regarded as an ancestor of soft wheats (3). We think that such a plant will never be found since it does not exist. Soft wheat is a product of evolution. In the light of cytogenetic studies the origin of soft

Fig. 3. *1*—the pollen einkorn plant, left; *2*—the amphidiploid, center; *3*—the maternal plant of *Tr. durum*, right.

wheats is regarded to have proceeded through hybridization of the 14-chromosome wheats with 28-chromosome ones followed by doubling of the chromosome number. In other words, soft wheat is a spontaneous amphidiploid which arose in nature long ago but considerably later than the 14 and 28-chromosome wheats did arise. This conclusion, we think, is rather clearly supported by the present investigation. The way indicated, it is true, is not the only possible mode of origin of the 42-chromosome wheats. Theoretically the 42-chromosome wheat may also be produced as a result of hybridization of the 28-chromosome species possessing genetically distinct genomes. This problem is being worked upon by the staff of this Department under my guidance.

To A. S. Afanassieva, A. B. Belova, M. V. Vedenov, A. A. Fedorov, M. G. Tom and other workers of the Genetic Department of the Timiriazev Agricultural Academy who helped me in carrying out the tests I would fain offer my heartiest thanks.

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GENETICS

PRODUCTION OF AMPHIDIPLOIDS OF *TR. DURUM* × *TR. TIMOPHEEVI*

By A. R. ZHEBRAK

(Communicated by N. I. Vavilov, Member of the Academy, 12. VI. 1939)

All the species of cultivated and wild wheat known to the present time fall into the following basic groups: hexaploid ($2n=42$), tetraploid ($2n=28$) and diploid ($2n=14$).

Species belonging to a common group which is characterized by the same chromosome number rather easily intercross and produce, as a rule, fertile progeny. An exception to this rule is offered by *Triticum Timopheevi* and *Tr. dicoccoides* which not only produce no fertile hybrids with species belonging to other chromosomal groups, but even with those of their own group. The hybrids are either completely sterile or produce a few seeds which are too sparse to build upon the valuable properties of the two species in question. Of a special value for hybridization is *Tr. Timopheevi* on account of its high tillering power, owing to which it suffers neither from the frit-fly nor the Hessian fly and is not attacked by fungous diseases. As such *Tr. Timopheevi* cannot be introduced into cultivation owing to its fragile spike and the difficulties experienced on its thrashing. It is therefore only through hybridization that its valuable properties can be made use of. However, as hybrids of the first and of the subsequent generations are nearly completely sterile, valuable properties of *Tr. Timopheevi* could not be used in synthetic plant-breeding. The problem of application of *Tr. Timopheevi* to synthetic breeding can be solved solely by surmounting the sterility of the hybrids of *Tr. Timopheevi* with other species.

Genetics possesses since long ago a method of experimental restoration of fertility of sterile hybrids by means of doubling the chromosome number. The discovery by Blakeslee and Avery (1) of the specific effect of colchicine upon doubling of the chromosome number in plants has attracted our attention to this kind of treatment. We attempted previously during several years to produce amphidiploids in *Tr. vulgare* with *Tr. durum* and *Tr. Timopheevi* by means of temperature treatment following the method of Dorzey (2). As no desired results were obtained by temperature treatment, we naturally decided to try the effect of colchicine upon wheat. We had to find dosages as no data on the crop in question were presented by the writers cited above.

After a number of preliminary trials we have chosen a concentration of 0.1%; in such a solution hybrid seeds were soaked during 24 hours being first treated with formalin and rinsed in water. We were in possession of

a small amount of kernels of the hybrid between *Tr. durum* v. *leucurum* and *Tr. Timopheevi* and have subjected to colchicine treatment only 8 hybrid kernels of this combination.

Seeds were placed into small glass vials with ground-in stoppers and after 24 hours soaking in colchicine solution without rinsing were placed into Petri dishes containing wet quartz sand. There they remained until germination, after which they were pricked off into pots and grew there until the development of roots suitable for fixation. After fixation the plants were transplanted into Mitcherlich's pots. From the 8 treated hybrid kernels *Tr. durum* × *Tr. Timopheevi* only 3 plants were obtained. One of them developed no kernels and differed in no way from the 10 control plants which were highly sterile; the second plant developed one kernel, whereas the third one produced three ears which contained seeds and the other three without seeds. The occurrence of a chimeral plant which bore highly fertile ears along with

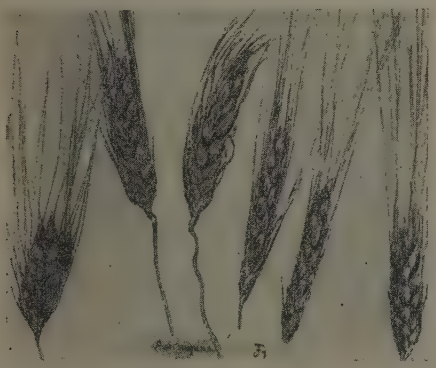

Fig. 1. Ears of the mother plants, hybrids F_1 (amphihaploid) and amphidiploids F_1: 1st ear—*Tr. Timopheevi*; 2nd and 3rd—amphidiploid; 4th and 5th—F_1, 6th—*Tr. durum*.

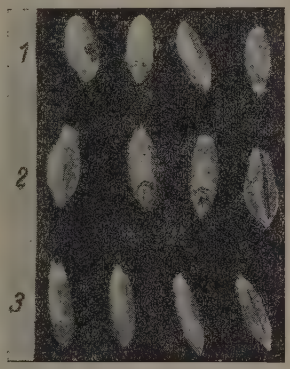

Fig. 2. Seeds: *1—Tr. durum*; *2*—amphidiploid *Tr. durum Tr. Timopheevi*, *3—Tr. Timopheevi*.

sterile ones witness for amphidiploidy. Fertility of the fertile ears was characterized by the following figures: in 14 spikelets of the first ear there were 30 kernels, in the second ear there were 26 kernels and in the 3rd one—15 kernels (Fig. 1). In the 85 ears of the control plants which were not subjected to colchicine treatment there were 1131 spikelets containing only one kernel.

The description of the characters of the amphidiploid's ear is made using material from the chimeral plant. It is possible that subsequent corrections with regard to some characters will be needed.

Coloration of the ear is white. Length equals 7 cm. Awns are adpressed and attain 9—10 cm in length. Kernels are tightly enclosed into the glumae but may be easily taken out with the forceps without injury to the ear. Spikelets attain 1.5—1.6 cm in length and are arranged in a straight line. The number of flowers per spikelet is 3. The total number of spikelets per ear is 16; two of them located at the base are sterile. The number of well filled kernels in 14 spikelets is equal to 30. More than 2 kernels fall thus upon individual spikelet. The density of the ear is equal to 2.3. The length of the spikelet glumae including dent is 10.5 mm. Its width equals 4 mm. The lateral nerve of the gluma forms a protrusion, the brachium of the gluma is acute. The length of the dent is 2 mm. The shape of the kernel approaches ellipsoidal one. The length attains 8—10 mm; the width, 4 mm. The embryo occupies about one fourth of the surface of the dorsal side of the kernel. Hairs

of the brush attain 1 mm. The total weight of the 30 kernels is equal to 1.98 g, that of the 26 kernels of the second ear 1.755 and the 15 kernels of the third ear weigh 0.86 g. Calculation gives thus a weight of 65 g per 1000 kernels.

The shape of the kernel is asymmetrical, one of its sides being smaller than the other. The coloration of the kernel is dark-red, darker than in *Tr. Timopheevi*. *Tr. durum* v. *leucurum* with white corn served in our cross as mother form. The character of the colour was thus transmitted from *Tr. Timopheevi* accompanied by a certain increase in pigmentation (Fig. 2).

Cytological investigation was conducted on the control plants, on chimeral plants and on 2 plants obtained from seeds from the first and second ears. The hybrid control plant of the first generations *Tr. durum*×*Tr. Timopheevi* had $2n = 28$ chromosomes. The same chromosome number was present in the rootlets taken at the 4—5 leaf stage from the chimeral plant which produced amphidiploid ears. In the rootlets of the 4 plants from seeds of the fertile ears there were 56 chromosomes (Fig. 3). The investigation has demonstrated the production of an entirely new wheat species ($2n=56$) in which the basic number 7 was repeated 8 times. This species belongs to the octoploid group which will be extended.

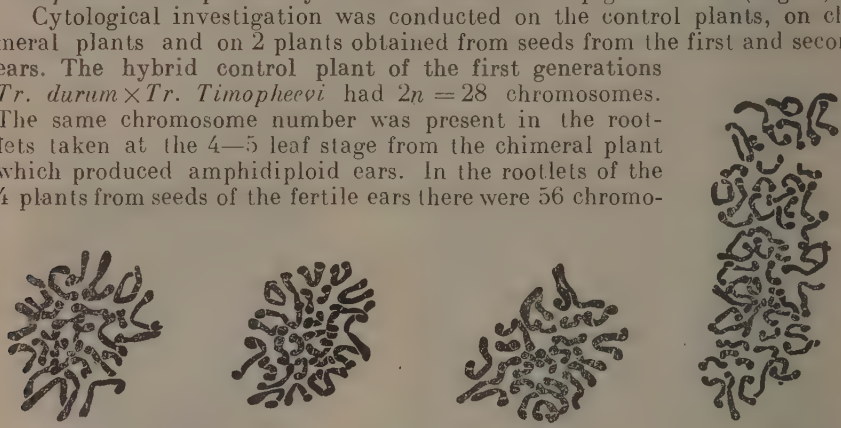

Fig. 3. Cytological slides: 1) *Tr. durum*; 2) *Tr. Timopheevi*; 3) F_1 *Tr. durum*×*Tr. Timopheevi* (amphihaploid); 4) F_2 amphidiploid *Tr. durum*×*Tr. Timopheevi* (*Triticum Sovetica*).

Produced by us at the Department of Genetics of the Timiriazev Agricultural Academy experimentally this species has been given the name *Tr. Sovetica*. The amphidiploid forms differ cytologically from the original species not merely in the chromosome number but in cell size, size of the nuclei, number of the nucleoli, etc. The cells and nuclei are larger in the amphidiploid and the number of nucleoli attains 6—8.

The following sizes in microns were determined from measurements of 100 cells and nuclei each belonging to the second layer of cells from 4 portions of a rootlet and 9 subsequent sections:

		Cells	Nuclei
Amphidiploid	greater diameter	30.2	18.8
	smaller »	22.3	15.2
Tr. durum	greater »	24.8	13.3
	smaller »	20.5	11.9
Tr. Timopheevi	greater »	25.4	13.2
	smaller »	19.4	11.9

The discovered difference in chromosome numbers of the rootlets and upper parts demonstrates that it is impossible to judge with assurance from cytological investigation of the rootlets alone when seeds have been treated by some active substance.

As to the character of development of the amphidiploids as compared with the original species one should note that they had thicker first leaves of an intense green colour. The length and width of the leaves of amphidiploids were also very different. The stem and the leaves were coarser.

The pubescence on leaves typical of *Tr. Timopheevi* is considerably weaker in amphidiploid. This character is inherited in an intermediate form. It was impossible to establish any connexion between the degree of development and the species which the maternal plant belonged to. The amphidiploids obtained on *Tr. durum* mother were similar in pubescence to the first generation obtained on *Tr. Timopheevi* mother.

Such a character as tillering is probably also inherited intermediately. *Tr. Timopheevi* tillers very strongly. The amphidiploid tillers stronger than *Tr. durum* but less than *Tr. Timopheevi*. When amphidiploids from *Tr. durum* mothers are compared with the first generation hybrids from *Tr. Timopheevi* mothers, an impression is gained of a stronger tillering of the hybrids from *Tr. Timopheevi* mothers. Conclusive evidence for this character will be available toward the end of the season after the counts of the stem number will be completed. As to immunity of the amphidiploids one may judge only on the relation to *Erysiphe graminis* of the first generation hybrid. Under similar conditions when grown in the greenhouse *Tr. durum* was rather strongly attacked by *Erysiphe graminis* whereas *Tr. Timopheevi* remained intact. The first generation hybrid was as resistant. Inasmuch as the amphidiploid possesses the same chromosal relations as the first generation hybrid, one should expect it to possess the same physiological properties.

It should be noted in concluding that colchicine is a very efficient agent for producing polyploidy. It may be used with success for the production of wheat *Agropyrum* and other amphidiploids.

As with pure species so with hybrid seeds the most frequent result of treatment are chimeral plants with sectors containing different chromosome numbers and producing shoots of different genetic constitution. The amphidiploid *Tr. durum*×*Tr. Timopheevi* obtained by us appears to be of great interest for practical breeding, since it makes it possible to use the valuable properties of *Tr. Timopheevi* for practical plant breeding. As to the question on the immediate practical value of the amphidiploid we are inclined to cautiousness since further study is necessary here.

Thanks are due to the collaborators of the Department of Genetics, A. S. Afanassieva, A. D. Belova, M. G. Tom and others for their co-operation in the experimental part of the work.

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GENETICS

GENETIC ANALYSIS OF ALBINISM MUTATIONS INDUCED IN BARLEY BY X-RADIATION

By I. N. SHALYGIN

(Communicated by N. I. Vavilov, Member of the Academy, 21. VII. 1939)

The nature of inheritance of spontaneous albinism mutations in barley was studied in detail by Nilsson-Ehle ([1], [2]) and Hallquist ([3], [4]) who concluded that there were several genes which in a homozygous state produce the same phenotypic effect. From this point of view experiments on the artificial induction of albinism mutations in barley by X-radiation of the germ cells and of dry and germinating seeds ([5], [6], [7]) are of interest.

As material for our investigation we used six families of two-rowed barley (*Hordeum distichum* var. *Steudeli*, Körn.), the progeny of which was characterized by a considerable percentage of mutations when treated with X-rays. We obtained this material through the kindness of A. N. Lutkov. Data on the behaviour in X_2 of each of these families heterozygous as regards albinism mutations are given in Table 1.

Table 1

Number of Albino Seedlings in X_2 of *Hordeum distichum* var. *Steudeli* Körn.

Family No.	Treated	X-ray dosage (in *r* units)	No. of seedlings in X_2		
			Total	Green	Albino
432 (1 head)	Germinating seeds	600	18	17	1
2643	Pollen	500	40	37	3
2595	»	500	33	28	5
2564	»	1500	22	19	3
1950 (1 head)	Germinating seeds	800	17	16	1
2118 (1 head)	Do	900	16	15	1

Counts of green and albino seedlings in X_4 confirmed the data previously obtained as to the monohybrid nature of the inheritance of albinism (Table 2). For convenience in handling the data, the progeny of each family, in the X_2 of which there appeared albino segregates, was given a group number. Thus, we designated the progeny of Family No. 432 as Group I, of Family No. 2643 as Group II, of Family 2595 as Group III, of Family No. 2564 as Group IV, of Family No. 1950 as Group V, and of Family No. 2118 as Group VI.

Table 2

Character of Segregation for Normal and Albino Seedlings in the X_4 of Heterozygous Families

Progeny groups used in cycle of crosses	No. of seedlings in X_4			Expected Nos. (3 : 1)	m	D/m
	Total	Green	Albino			
I	1441	1112	329	1080.75 : 360.25	±15.93	1.96
II	585	454	131	438.75 : 146.25	±10.08	1.51
III	698	536	162	523.5 : 174.5	±11.15	1.12
IV	963	738	225	722.25 : 240.75	±13.13	1.20
V	672	520	152	504 : 168	±10.84	1.48
VI	156	120	36	117 : 39	± 5.26	0.57

In our cycle of crosses we included all 15 possible combinations. About 500 crosses were made of each combination. On each of the parental plants two or three control heads were left for self-pollination. After the barley ripened there were first sown only seeds from the self-pollinated control heads, which made it possible to choose pairs of plants heterozygous as regards albinism.

All the seedlings of the first generation, totaling more than 600 plants, proved to be green. In the second generation there were 88 homozygous families, 172 families segregating in a 3 : 1 monohybrid ratio, and 75 families segregating into green and albino plants in a 9 : 7 ratio. On the basis of the character of segregation in F_2 and the absence of albino plants in F_1, the only possible conclusion to be drawn is that the genes for albinism in each of the six groups investigated belong to six different allelomorphic pairs. For each pair of heterozygous plants crossed this may be represented in the following general form:

Parents: $A_1a_1 \quad A_2A_2 \times A_1A_1 \quad A_2a_2$

Gametes: $A_1A_2, \quad a_1A_2; \quad A_1A_2, \quad A_1a_2$

F_1 genotypes: $A_1A_1A_2A_2, \quad A_1a_1A_2A_2, \quad A_1A_1A_2a_2, \quad A_1a_1A_2a_2$

From the above scheme of inheritance it is to be expected that in the F_2 an average of 25% of the families should be homozygous, 50% of them heterozygous with a monohybrid type of segregation, and 25% of them heterozygous with a segregation in a 9 : 7 ratio. A comparison of the expected figures (83.75 : 167.50 : 83.75) with those actually obtained (88 : 172 : 75) shows that they agree quite closely.

In Table 3 are given the data on the segregation of the 172 heterozygous families for each separate combination.

The data characterizing the mono- and dihybrid segregations of the heterozygous families (Tables 2, 3 and 4) show, as a rule, in each combination a somewhat smaller number of albinos than expected. Presumably, the lethal nature of albinism in a homozygous state is, to some extent, manifested even prior to the seedling stage. While in case the germinating power of seeds of plants heterozygous and homozygous for albinism is the same, an inconsiderable shortage in some families of one type of seedling is compensated for by a corresponding surplus of this type in other families, in case of an uncompensated shortage of albinos in all families, with an increase in the number of investigated individuals, the ratio D/m will not decrease but will have a tendency to increase.

Very few crosses were made of the combinations I×V and III×V, which probably accounts for the absence in the F_2 from these combinations of fa-

Table 3

Segregation for Normal and Albino Seedlings in the F_2 of Heterozygous Families Carrying One Gene for Albinism

Groups included in the crosses between heterozygous families	No. of seedlings in F_2			Expected Nos. (3 : 1)	m	D/m	No. of families
	Total	Green	Albino				
I×II	385	299	86	288.75 : 96.25	± 8.17	1.25	10
I×III	506	390	116	379.5 : 126.5	± 9.45	1.11	14
I×IV	233	180	53	174.75 : 58.25	± 6.40	0.82	6
I×V	541	414	127	405.75 : 135.25	± 9.86	0.84	7
I×VI	727	557	170	545.25 : 181.75	±11.41	1.03	12
II×III	271	209	62	203.25 : 67.75	± 6.91	0.83	7
II×IV	468	360	108	351 : 117	± 9.11	0.99	10
II×V	735	566	169	551.25 : 183.75	±11.41	1.29	12
II×VI	1 117	860	257	837.75 : 279.25	±14.07	1.59	21
III×IV	1 535	1 178	357	1 151.25 : 383.75	±16.55	1.62	29
III×V	331	253	78	248.25 : 82.75	± 7.72	0.61	4
III×VI	209	161	48	156.75 : 52.25	± 6.08	0.70	6
IV×V	763	584	179	572.25 : 190.75	±11.71	1.00	11
IV×VI	908	698	210	681 : 227	±12.71	1.34	17
V×VI	216	166	50	162 : 54	± 6.19	0.65	6

Table 4

Segregation for Normal and Albino Seedlings in the F_2 of Heterozygous Families Carrying Two Genes for Albinism

Groups included in the crosses between heterozygous families	No. of seedlings in F_2			Expected Nos. (9 : 7)	m	D/m	No. of families
	Total	Green	Albino				
I×II	79	48	31	44.4 : 36.6	± 4.34	0.83	3
I×III	128	76	52	72 : 56	± 5.60	0.71	5
I×IV	104	60	44	58.5 : 45.5	± 5.04	0.30	3
I×VI	453	272	181	254.8 : 198.2	±10.42	1.65	10
II×III	158	93	65	88.88 : 69.12	± 6.18	0.67	5
II×IV	174	106	68	97.88 : 76.12	± 6.44	1.26	4
II×V	80	49	31	45 : 35	± 4.36	0.92	2
II×VI	400	220	180	225 : 175	± 9.95	0.50	10
III×IV	572	341	231	321.75 : 250.25	±11.73	1.64	13
III×VI	146	90	56	82.13 : 67.87	± 5.87	1.34	4
IV×V	388	236	152	218.25 : 169.75	± 9.62	1.84	7
IV×VI	464	264	200	261 : 203	±10.67	0.28	7
V×VI	95	59	36	53.45 : 41.55	± 4.75	1.17	2

milies segregating in a 9 : 7 ratio. Nevertheless, it is clear that the gene for albinism in Group V can be identical only with the gene for albinism in one of the two groups, I or III. It is very probable that a number of new gene loci inducing albinism in barley may be established. In some crop plants that have been studied in greater detail in this respect there are already known a large number of phenotypically indistinguishable mutations induced by different genes (in *Zea Mays*, for example, there are known to be 11 genes for albinism).

The normal chlorophyll colour depends on many genes, each of which plays a positive rôle only in combination with the others. If we represent the

dependence of chlorophyll formation on the genotype as $A_1A_1A_2A_2A_5A_5\ldots A_n\,A_n$, it must be presumed that all members of this allelomorphic series are linked with one another in determining chlorophyll pigmentation. It suffices to include in the chain of harmoniously acting genes a homozygous pair of qualitatively altered genes, and the entire system is so disturbed that the formation of the normal chlorophyll colour is no longer possible.

Hence, the frequency of occurrence of albinism mutations not only depends on the greater or lesser degree of gene mutability, but may also depend on the presence of many genes, each of which mutating under the effect of X-radiation gives a phenotypic effect indistinguishable from that produced by any of the others. In such cases a genetic analysis helps to ascertain the actual mutation rate of genes.

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GENETICS

ANALYSIS OF VITALITY OF FLIES OF *DROSOPHILA MELANOGASTER* HETEROZYGOUS FOR LETHALS ARISEN IN NATURE

By R. A. MASING

(Communicated by N. I. Vavilov, Member of the Academy, 21. VII. 1939)

While conducting cytogenetic analysis of lethals arisen in flies of *Drosophila melanogaster* under natural conditions we (1) have discovered cases when vitality* of the flies heterozygous for the lethals was increased as compared with those containing no lethals. These facts, however, were revealed accidentally. It was of interest to conduct a systematic study of the relative vitality of the heterozygotes for lethals in some of the natural populations in order to see how often such a phenomenon occurs and to determine whether or not decreased vitality, i. e. semidominant manifestation of lethals, takes place.

In view of this goal flies of *Drosophila melanogaster* of the Crimea (Nikita Botanical Garden) population were taken in 1937.

The lethals in the II chromosome were found by the usual $\frac{Cy}{L_2}$ method. While examining the F_3 cultures for lethals, we were especially looking for genes for increased vitality, but failed to find a single case of this sort among 150 chromosomes**. In the given population the proportion of lethals proved to be high (22%). All the lethals were balanced with the aid of the *Cy* line.

Comparative vitality of flies heterozygous for lethals was studied in 32 lines. In order to analyze vitality males $\frac{Cy}{l}$ from the line investigated were crossed, as in the previous work, with females from normal and ebony lines which were inbred during a long time. Virgin females were subjected to inbreeding. Several individual cultures were started in each generation and the best culture was selected for perpetuation of the line; thus, selection for good vitality was effected.

From crossing $\frac{Cy}{l}$ males with normal females were obtained flies $\frac{Cy}{l}\frac{+}{+}$ and $\frac{+}{l}\frac{+}{+}$; from the cross with ebony were produced flies $\frac{Cy}{+}\frac{+}{e}$ and $\frac{+}{l}\frac{+}{e}$. As a control served crossing of males $\frac{Cy}{+}$ with normal and ebony flies from the same inbred lines. In all the crosses there was expected a 1 : 1 phenotypic ratio of *Cy* flies to normal ones. Deviation from this ratio in the expe-

* Under vitality we understand vitality during development, not vitality of adult flies.

** A gene for increase vitality was described in the sex-chromosome by Timofeeff-Ressovsky (4). It was produced in experiments on X-raying flies.

rimental cultures should indicate instances of increased or decreased vitality of the flies heterozygous for lethals*.

Table 1

Cross		% of normal flies			
No. of the culture of males $\frac{Cy}{l}$	Females	Observed	Expected	$\frac{\text{Diff}}{m}$	n
505	ebony	24.1	50±1.53	16.9	1 075
	normal	32.3	50±2.08	8.5	579
627	ebony	58.2	50±1.75	4.7	818
	normal	59.9	50±2.31	4.3	466
513	ebony	41.7	50±1.77	4.7	793
	normal	56.8	50±1.26	5.4	1 576
614	ebony	41.9	50±1.19	6.8	1 767
	normal	53.9	50±2.22	1.8	510
518	ebony	61.5	50±0.88	13.0	3 215
	normal	47.4	50±1.35	1.9	1 370
Control	ebony	51.7	50±1.31	1.3	1 461
	normal	50.9	50±1.40	0.6	1 267

Cultures were conducted in Erlenmeyer's flasks under strictly identical external conditions. Five females and 5 males were placed into each flask and after 4 days the parents were removed. Flies were counted in each flask separately; in case different flasks yielded divergent results, no averages were computed but the culture in question was regarded as an uncertain case. In each culture flies were counted till the end of the summer, as it was noted that in certain lines normal flies hatch earlier than *Cy* flies, so that the result could be distorted in case counting were finished too early.

In Table 1 are presented the data on the analysis of the vitality for all the most characteristic lines. One can see very different vitality of the heterozygotes for the lethals in different lines. It is shown in Table 2 how many lines of each type of vitality (if one may say so) were met with in the lines investigated.

Table 2

Cross of males $\frac{Cy}{l}$ with females		Number of the cultures
ebony	normal	
—	—	8
—	none	6
—	+	1
—	0	1
+	+	2
+	0	2
+	+	6
?	?	6
	Total	32

Note. Sign — denotes decreased vitality; sign + an increased one; 0 — normal vitality (coincidence of observation with expectation); ? — vitality fluctuated in the one or the other direction within the same experiment in different flasks or in repeated experiments.

In 16 lines out of 32 there was observed a decreased vitality of flies heterozygous for lethals, if one includes into this group lines showing diminished vitality with at least one single line out of those involved in the crosses. This is in opposition to the observations by Sturtevant [3] who has demonstrated full recessiveness of the lethals which arose under natural conditions in *Drosophila pseudoobscura*. In 4 lines there was observed an increased vitality of the heterozygotes for lethals; only two of them showed increased

* The chromosome bearing *Cy* is also lethal, but inasmuch as it was present in all the investigated lines its presence did not interfere with the analysis.

vitality when crossed with the two remaining lines. In 3 lines there was observed a peculiar strong difference in the indices of vitality according as the cross was made with normal or ebony flies; this phenomenon was noted in the preceding investigation. Conflicting evidence was secured for 6 lines, either in different flasks or when experiments were repeated; finally, in 6 cases observation coincided with expectation.

All the 32 lethals were identified. Identical lethals proved to be scarce. One of the lethals was found 3 times, the other 2 times. It is of interest to note that identical lethals gave different indices of vitality of the heterozygotes for the lethal. In the cases in question it is not merely the lethals which appear to affect vitality of the heterozygotes. However, without an analysis of the lines concerned for individual regions of the chromosome containing the lethal and that lacking it, no definite conclusions could be made.

It seems to be of extreme interest that in a number of the lines studied (in 20 out of 32) heterozygous condition with regard to the lethal affects in the one or the other manner vitality of the individuals.

However, all the phenomena found by us were produced by crossing males $\frac{Cy}{l}$ from the line studied with females from laboratory lines; a question arises whether the same results will be produced when males $\frac{Cy}{l}$ are crossed with females of the same Crimea population.

In order to clear up this question we took again in 1938 lethals of the II chromosome of the same Crimea population which were obtained and kindly handed over to us by R. L. Berg. Males $\frac{Cy}{l}$ from the lines studied were crossed with females from their respective population. For the sake of control identical females were crossed with males $\frac{Cy}{\text{Nik. Garden}}$. In the case in question the control exhibited considerable predominance of Cy flies over normals (Table 3).

Table 3

Crossing of Males $\frac{Cy}{l}$ with Females of the Crimea Population

No. of the culture of the males $\frac{Cy}{l}$	% of normal flies		Diff $\frac{}{\sqrt{m_1^2 + m_2^2}}$	n
	Observed	Expected		
63	33.6±1.71	42.8±0.69	4.9	758
155	48.5±1.33		4.4	1926
11	37.7±1.42		3.2	1160
129	46.8±1.58		2.3	992
Control	42.8	50±0.69	10.4	5188

In Table 3 are presented the results of the study of vitality of heterozygotes for the lethals in the most characteristic lines obtained in 1938. In Table 4 prepared in a manner similar to that of Table 2 is shown how many lines of each type of vitality were found in the 26 lines investigated.

In 8 lines heterozygotes for the lethal were characterized by a lowered vitality and only in 2 cases by an increased one. In a comparatively high proportion of lines (in 15 out of 26) the vitality of the heterozygotes was identical to that in the control.

An attempt was made to conduct an analysis of vitality by means of crossing with females of the same normal inbred lines used for analysis in 1937; this was done parallel to crossing with females of the population

which males belonged to. However, in the normal line in question during 4 months neither inbreeding nor selection for good vitality were conducted, and so many lethals* had accumulated in it that it proved to be entirely unsuitable for analysis.

Table 4

Index of vitality	Number of cultures
—	8
+	2
0	15
?	1
Total: . .	26

Note. Designations are the same as in Table 2.

In the main, crossing males $\frac{Cy}{l}$ with females of their own population gave the same result as crossing with the laboratory lines. We again, and this is of special importance, were confronted with cases of both increased and diminished vitality of heterozygotes for lethals.

The phenomenon of diminished vitality of the heterozygotes for lethals finds its good explanation in the semidominant expression of the lethals. More difficult is the explanation of the increased vitality except for cases of dominant genes of increased vitality with recessive lethal effect which probably sometimes occur (2). It may possibly depend on accumulation of modifiers in the process of evolution which not only neutralize the deleterious effect of the lethal, but even lead to increase of vitality of the heterozygotes. In the line *T*-201 studied in detail by us (2) increased vitality was depending on interaction between a lethal and a modifier.

Preliminary experiments on comparative vitality of adult flies of the two groups studied have demonstrated independence of this sort of vitality of that shown in the course of development. Probably, the struggle for life between $\frac{Cy}{+}$ and $\frac{l}{+}$ occurs at early stages of development. The phenomenon of increased vitality was first discovered in the study of lethals with the aid of Painter's method. In order to obtain larvae heterozygous for lethals males $\frac{Cy}{l}$ were crossed with ebony flies. In some of the cultures a very low number of *Cy* larvae was apparent.

Attention was attracted to the fact that in the majority of lines when cultures $\frac{Cy}{l} \times \frac{Cy}{l}$ were conducted during a long period of time both the increased and decreased vitality of the heterozygotes for the lethals disappear. In view of this fact the analysis of the vitality of the heterozygotes for the lethals should be conducted immediately upon obtaining lethals from the nature.

We may see at any event that lethals in heterozygous state are often not indifferent for the individual's life. Inasmuch as in 3 lines (2) we have demonstrated a connexion between increased vitality of flies and the presence of a lethal in heterozygous state, we have still more grounds to suppose that selection in heterozygous state plays an important part in dynamics of populations; in respect to lethals it may proceed both in plus- and minus-direction but the latter case is the more common one.

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* Lethals were discovered with the aid of a special investigation.

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CYTOGENETICS

MEIOSIS IN INBRED RYE

By N. T. KAKHIDZE

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According to a number of writers the specificity of inbreeding genetically considered consists in the fact that the average hybrid population of a cross pollinating plant when selfed segregates into diverse forms differing in a number of characters, especially in fertility; this wide variability is regarded to be connected with accumulation of certain genes in homozygous conditions, these genes causing depression or stimulating the development of plants in the progeny from inbreeding([1],[2],[3]). Inasmuch as genetic constitution of the organism is reflected in the course of the reduction division, it should be expected that inbreeding will show a definite cytological effect, too.

Meiosis was studied by me in inbred forms of various lines of inbred progenies in the rye of variety «Vjatka». Material was obtained from the section for rye of the All-Union Institute of Plant Industry. Thanks are due to V. E. Antropova for this material.

In the analysis of meiosis the main attention was concentrated upon the comparison of the various inbred forms with respect to the presence of bivalents and to the average frequency of chiasmata as a criterion of the degree of conjugation ([4]). Chiasmata were counted in all the inbred plants studied by me at the metaphase stage, and in some of them at diakinesis, too. All the individuals analyzed had $2n=14$. The number of chiasmata per bivalent varied from zero to four and the types of the bivalent structure were the same as that observed in ordinary forms. However, the frequency of occurrence of this or that type differed from the one found in normal forms; here bivalents with one chiasma as well as univalents were much more frequent which accounted for a resulting decrease in the mean value (table).

«Vjatka»	Metaphase		Diakinesis	
	Average number of chiasmata	Terminalization coefficient	Average number of chiasmata	Terminalization coefficient
Norm	2.23 ±0.027	0.71	2.21 ±0.042	0.70
J_6 1809	2.11 ±0.084	0.74	2.13 ±0.015	0.70
J_8 17	2.01 ±0.019	0.71	2.108±0.9	0.68
J_8 738	1.957±0.092	0.56	—	—
J_8 404	1.67 ±0.127	0.62	1.65 ±0.056	0.66
J_6 45	1.175±0.11	—	—	—

Comparing data on different individuals revealed differences among them, mainly in the frequency of this or that type of the bivalent. Plants J_8 1809 and J_8 17 had the average number of chiasmata and types of bivalent very closely approximating those in the norm (Fig. 1). In J_8 738 the average frequency of chiasmata was perceptibly lower than the normals one and the degree of conjugation was less. J_8 404 was characterized by a considerably lowered average frequency of chiasmata and by preponderance of bivalents possessing one chiasma (Fig. 2). Lagging of the one of the bivalents on the equator was also observed. J_6 45 has shown the strongest deviations. They were expressed not only in a very low average number of chiasmata, but also in more significant irregularities; not infrequently univalents were found at metaphase, their number varying from zero to six (Fig. 3); lagging of the chromosomes at the first anaphase (Fig. 4) and bridges between the poles were observed (Fig. 5) and formation of fragments probably connected with the presence of bridges. Remnants of bridges could be sometimes detected even at the close of the second meiotic division.

The behaviour of the chromosomes in inbred rye was also considered in the work by Lamm ([5]) which is the only one to analyze the cytology of in

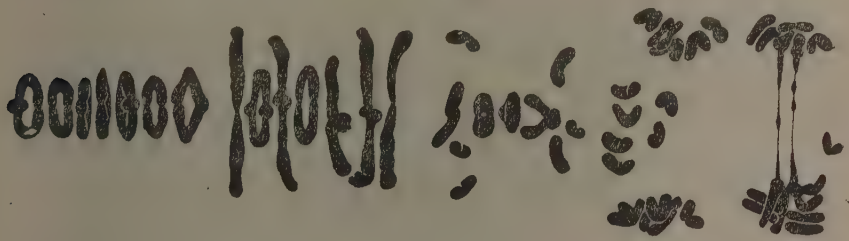

Fig. 1. Fig. 2. Fig. 3. Fig. 4. Fig. 5.

bred plants. Our study which was conducted independently of the observations by Lamm gave results which closely approached those contained in his paper cited above.

The results of the two investigations thus point to the fact that as a result of inbreeding individuals are isolated which differ from one another in the course of the process of meiosis, mainly in the formation of bivalents.

What are the possible reasons for such a phenomenon?

The morphological condition of bivalent formation is in the first line a certain degree of compatibility of the chromosomes entering conjugation. The possibility of chiasma formation is determined by the fact that at a given point homologous parts of the chromosome conjugate, i. e. the chiasma frequency is connected with the degree of homology of the partners which compose the bivalent. In the light of this consideration very instructive is the fact that the mean frequencies of chiasmata in hybrids may be considerably lower than in the parents ([6]), whereas close affinity of the forms crossed is favourable to a higher chiasma frequency ([7]). On the other hand, structural changes may disturb conjugation, the constitution of the hereditary material remaining the same ([8]). From a genetic point of view the action of genotypic control is considered; it consists in the existence of genes causing variation in the number of chiasmata, irregular separation in meiosis, etc. ([9, 10, 11, 4]). Besides, the data by Sax ([12]) on comparison of the average frequency of chiasmata in different geographic races and under different conditions showed, that while being a genotypic character the chiasma frequency is subject at the same time to external influences which may be met with in nature.

While analyzing morphologically the bivalent formation in inbreeding one should bear in mind, on the one hand, that the degree of similarity of the homologous chromosomes should increase from close inbreeding. At the same time in the individuals of the inbred progeny the bridges at the first anaphase described above are observed much more frequently than in the norm; and formation of bridges is connected with structural changes of the chromosomes ([13,5]). Their presence inspite of inbreeding in a homozygous condition may be explained either by their origination in the inbred progeny under the influence of certain genotypic combinations ([14]) or as a result of a general depression of the plant. However, the structural changes alone are insufficient for explanation of such striking disturbances as those found, for example, in the plant J_6 45. Direct action of genotypic control in connexion with a general depression in such plants seem to be a satisfactory explanation of the phenomenon in question.

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CYTOGENETICS

NEW POLYPLOIDY-INDUCING CHEMICALS

By M. V. FAVORSKY

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All authors ([1—10]) who have studied the substances inducing polyploidy explain the mechanism of their action upon cell division by disturbances in the achromatic spindle. These substances belong to very different groups of carbon compounds. I therefore was not inclined to suppose that the only factor in the specific action of the colchicine is the polycyclic structure of its molecule, but chose for my first experiments the dyestuff aurantia, a derivative of diphenylamine which is known to produce cutaneous disease in man if introduced under the skin, thus approaching to the colchicine in this particular.

My further way was to apply substances whose molecule was a similar but increasingly complex structure, with the special aim to ascertain the rôle of lateral groups as far as possible. Some of the tests were run with the colchicine and acenaphthene for the sake of comparison. At present 12 substances have been tested for their influence on mitosis—aurantia, diphenylamine, tribromoaniline, α-, β-naphthylamine, nitrobenzene, trinitrophenol, tribromophenol, phenol, iodoform and dichlorodiethylsulphide. Three of them (aurantia, $NH_4N[C_6H_2(NO_2)_3]_2$; diphenylamine, $NH[C_6H_5]_2$; tribromoaniline, $NH_2C_6H_2Br_3$) have called forth a profuse formation of polyploid cells in the roots of the experimental plants.

All those substances were tried on the same material—barley *Hordeum distichum* var. *nutans* Schübl., variety Wiener. The treatment started either at the outset of germination or when the roots had already emerged from the grain and attained a length of 1 cm. Its duration varied from 1 to 6 days. 9 four-day-old seedlings of barley treated with various substances are shown in Fig. 1 against the control plants. Rather easily soluble substances were applied as weak water solutions (colchicine—0.25 and 0.025%, aurantia—0.01 and 0.001%, aniline—0.1, 0.05 and 0.001%, trinitrophenol—0.1 and 0.01%, phenol—0.01%). The tribromoaniline, almost insoluble and non-volatile, was sprayed abundantly as powder. Sparingly soluble and poorly volatile substances as acenaphthene (diphenylamine, α- and β-naphthylamine, tribromophenol) and substances which pass off readily in the form of vapour were applied per se (iodoform 0.01 g per 200 cm³ of air) or as 1% solution in vaseline (nitrobenzene and dichlorodiethylsulphide, 1—2 g of solution per 200 cm³ of air). My endeavour while setting the dosage was to call forth an at least visible delay of growth without killing the plant. The fixative for the material of cytological investigation was largely S. Navashin's chrom-formol-acetic mixture, constant preparation being stained with fuchsin-sulphuric acid according to Velgen and with iron haematoxylin according to M. Heidenhain.

In my experiments the effect of treatment with aurantia, diphenylamine, tribromoaniline, nitrobenzene and iodoform was to produce root swelling in the region of elongation which was suggestive of that appearing in response to treatment with colchicine and acenaphthene (Fig. 1). If germinated seeds are acted upon by a 0.001% solution of aurantia or tribromoaniline, small swellings become usually distinct in 24 hours. They make further growth gradually and at 22°C reach a maximum after two or three days. The most swelled portion is three times the breadth of a normal root. As the tumour develops, the growth of roots is more and more feeble until it is checked completely. The filaments on the plan of swelling do not stop growth, however, and develop up to their normal length. The swelling is more and more close to the sheath. The meristematic zone, which on the second day is sharply separated by the tumour, becomes gradually shortened and fuses with the swelling so that nothing from it is left, as a rule, on the fifth or sixth day. Sometimes cracks appear on the tumour, it shrinks slightly and a few days

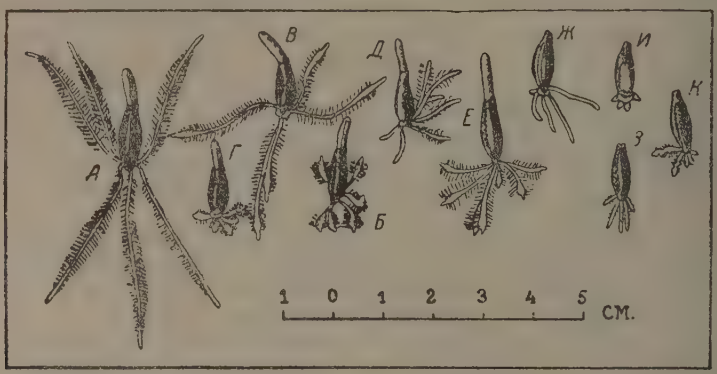

Fig. 1. Four-day-old seedlings of barley. *A*, control; *Б*, grown under the action of acenaphthene; *B*, treated with 0.025% solution of colchicine; *Г*, with 0.001% solution of aurantia: *Д*, with diphenylamine: *E*, tribromoaniline; *Ж*, 0.04% solution of aniline; *З*, α-naphthylamine: *И*, nitrobenzene vapour; *К*, iodoform vapour.

later the radicle perishes entirely. After the swelling has attained considerable dimensions, the radicles cannot be restored to normal life even by washing the plants and taking them into usual conditions of growth. In this case the tumour continues growing, the roots perish and new ones of a normal appearance form instead. But for the delayed development of the experimental seedlings they would differ in nothing from the controls. However, under the action of acenaphthene or even a mean (0.025%) solution of colchicine, the swelling of the coleoptile is poor and is not observed until the seedling emerges above the membranes. Such a solution of colchicine acting directly upon the roots causes the tumours to appear later than in the case of 0.001% aurantia, acenaphthene or tribromoaniline (Fig. 1).

The effect produced by diphenylamine is similar to that of aurantia and tribromoaniline. Growth is strongly inhibited, but the tumour starts developing later—usually after two or three days of treatment. The influence of nitrobenzene vapour upon germinated plants makes the zone of root elongation swell slightly as early as after 10—12 hours. Making further growth, however, the tumour fails to reach the size which is attained under the influence of the above substances. If the grains germinate in the presence of nitrobenzene vapour in the atmosphere, the coleorhiza and coleoptile swell when still under the membranes. The seedlings remind of those which appear under

the action of a 0.25% solution of colchicine. The treatment may have lasted considerable time, but if it is stopped when the roots are still alive, the tumours shrink slightly, growth is resumed and the normal appearance of plants is restored in the course of time. The iodoform induced delayed growth and poor swelling followed by cracking of the whole root except for its meristematic part. The effect of the other substances tried by the author was only to delay growth.

The aurantia, diphenylamine and tribromoaniline are so much similar to one another and to colchicine and acenaphthene with regard to the modifications they induce in the cells and roots tissues, that it is impossible to determine from the microscopic preparations which of them is responsible for the change. The chromosomes are strongly shortened during mitosis. When the late prophase is passing into metaphase, it is only the nuclear membrane and the nucleoli which disappear. The arrangement of chromosomes is not in the

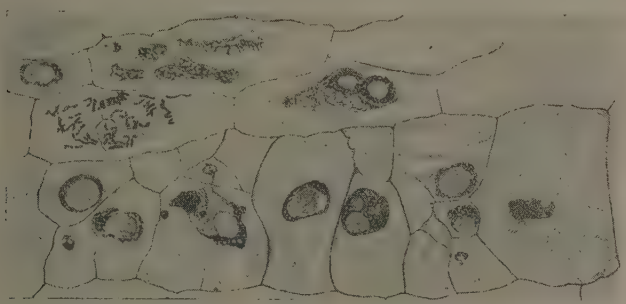

Fig. 2. Portion of tissue from a root tip of three-day-old barley seedling grown in the presence of tribromoaniline. Cells containing many chromosomes and nuclei along with citomeres of M. Navashin are visible.

equatorial plane. No achromatic spindle is formed. However, with the method used by the author the spindle was indistinct in the control preparations, too. The halves of the split chromosomes go then out of contact with each other, but not to the opposite poles. The anaphase differs from the metaphase in exhibiting a double number of chromosomes which are thinned and arranged in pairs. There are formed two or several daughter nuclei of no equal size in most of the cases.

Cell partitions do not form at all or else they divide the cell confusely in varying divisions. Citomeres of M. Navashin are formed (Fig. 2). This formation is easy to understand if one remembers that, according to the results obtained by Shakhlevich ([11]) with chloral hydrate, the achromatine apparatus is desorganized rather than annihilated. Similar formations were observed by Levan ([12]) under the action of colchicine. All division figures were of irregular shape. After two or three days of treatment most of the dividing cells are filled up with chromosomes spread confusely in great numbers. Their diploid and tetraploid sets are of seldom occurrence. The resting cells in the meristematic zone frequently contain a large irregularly formed nucleus or else there are several nuclei of varying form in each of those cells. Small chromatine clots were also observed. The cells having many chromosomes increase in dimensions. In the swollen part of the root the cells are strongly enlarged, though only one nucleus of normal size is contained in each. The mitoses cease after four or five days of treatment. Desorganization of the achromatine spindle and strong reduction of chromosomes result also from application of α-naphthylamine, but in this case these phenomena were observed on the

verge of cell's life and therefore we were not successful in following the mitosis until the anaphase. Most of the root cells were killed. Other new substances tested by the author did not desorganize the achromatine apparatus and failed to induce polyploids.

In the growing cones of the stems and in the leaves no polyploidy could be induced by any of the chemicals tried even after 5 days of treatment. The effect of colchicine, however, was to disturb mitosis in all parts of the plants in the same way as in the roots. 10 plants of barley treated with tribromoaniline on germination and 5 plants treated with 0.01% solution of aurantia during 5 days being then pricked off into pots in the open gave no tetraploid shoots. Of the first 30 new roots taken from those plants for cytological analysis only one root from a plant treated with aurantia proved to consist largely of tetraploid cells.

The changes induced in the roots by the substances here described are similar to those obtained by many authors with colchicine and acenaphthene. Root swelling is not always attended by desorganization of the achromatine apparatus and creation of polyploid cells. The colchicine favourably compares with other substances being able to penetrate the most safely protected parts of the plant. However, it seems probable that if an adequate method will be devised, the new substances tested can also be turned to account for obtaining polyploid plants.

The amino-group links up the new substances suggested with the colchicine. It is also contained in the aniline dyestuffs whose strong caryoclastic activity was pointed out by Dustin and co-workers ([13]) (tripanblau, isoaminblau, tripaflavin) and Politzer ([14]) (neutral red and auramine). The influence of these substances upon the mitosis in animals is suggestive of what is observed with substances inducing polyploidy in plants. If the amino-group in tribromoaniline is replaced by the hydroxyl (tribromophenol), the activity of the substance is entirely lost. Neither are the polyploid cells induced by trinitrophenol. The phenols are non active. This is also reported by Shmuck ([15]). The nitro-groups and halogens (bromine) introduced in the nucleus of the cyclic amine increase its activity. In the experiments by Shmuck ([9]) the bromine increased the activity of carbohydrates. From the experiments of other authors it will be obvious that polyploidy can be induced by substances, in which no amino-group (acenaphthene) nor even a cyclic structure of molecule (chloral hydrate, ether) is available. The organism is endowed with a constant response to widely varying stimulation.

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CYTOGENETICS

STRUCTURAL CHROMOSOME ALTERATIONS IN *TRITICUM DURUM*

By M. A. SIZOVA

(Communicated by N. I. Vavilov, Member of the Academy, 19. VII. 1939)

In the study of chromosome morphology of *Tr. durum expansum* v. *melanopus* and *Tr. durum abyssinicum* v. *gondaricum* there was found a clear difference between two chromosome pairs with appendices which are present in the sets of the two varieties. This difference consists in a strong

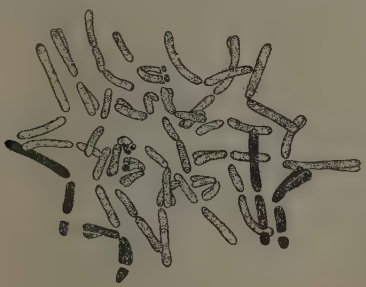

Fig. 1.

Fig. 2.

shortening of an arm in one of these pairs in the Abyssinian race (unpublished data by V. A. Poddubnaja-Arnoldi from the Cytological Laboratory of the All-Union Institute of Plant Industry).

To clear up the nature of this difference we studied the second hybrid generation from the cross *Tr. durum abyssinicum* v. *gondaricum* with. *Tr. durum expansum* v. *leucurum* from Sardinia; it was to some extent characterized by the same chromosome differences*.

First rootlets of 20 F_2-plants were investigated. Three groups of plants differing in chromosomes with appendices were established. The first group consisted of 10 plants (out of the 20 investigated) which possessed three strongly heterobrachial chromosomes with appendices and one small heterobrachial one (Fig. 1). The latter evidently represents the shortened chromosome of v. *gondaricum* whereas two of the remaining three belong to v. *leucurum*, the third indistinguishable from them being the longer chromosome of v. *gondaricum*.

The second group (5 plants) possessed chromosomes with appendices identical to those obtained in the set of v. *gondaricum*, i. e. two large strongly heterobrachial and two small isobrachial ones.

* The seeds of the hybrids were obtained from the Section for Wheat from . . Nikolaenko to whom my thanks are due.

The third group (3 plants) was similar in chromosomes with appendices to v. *leucurum*, i. e. all the chromosomes with appendices were large and strongly heterobrachial. In two remaining plants the morphology of the chromosomes remained unknown.

All the three groups of plants were studied with respect to reduction division. The material was fixed in chrome-acetic acid formalin 10 : 4 : 1; anthers were pretreated with a mixture of alcohol with acetic acid. The microtome sections were 18 microns thick and stained with iron haematoxylin.

In the first group, with a set asymmetrical in respect to the chromosomes with appendices, a complex of four chromosomes and 12 ordinary bivalents was found at diakinesis in all the pollen-mother-cells. Such a «quadrivalent» had in some cases the shape of a cross or of a ring (Fig. 2, *A*, *B*), in others of a twisted ring (Fig. 2, *C*), of an opened ring (Fig. 2, *D*). At the metaphase stage the «quadrivalent» either persisted in the shape of a ring (Fig. 2, *E*) or, in some cases, was twisted (Fig. 2, *F*). In investigating the reduction division in the two remaining groups of plants where no asymmetrical pairs of chromosomes with appendices were found, i. e. where they were of the type of v. *gondaricum* or v. *leucurum*, we have found in all the pollen-mother-cells only the 14 bivalents common in hard wheat.

The chromosomes of the «quadrivalent» are evidently homologous in some of their regions to two different chromosomes and a ring results from such homologous binding. Such data lead to a conclusion that formation of «quadrivalents» in the first group of plants is a result of reciprocal translocation which occurred in the one of the parental forms in the process of their evolutionary divergence.

The question which of the chromosomes participated in translocation is to some extent answered by the fact that the «quadrivalent» is always connected by two of its chromosomes with the nucleolus (Fig. 2, *A*, *B*, *C*, *D*, *E*, *F*). Special studies have shown(1) that in wheat just the chromosomes with appendices are connected with the nucleolus. On the basis of these data we can assert that two chromosomes with appendices participated in the «quadrivalent» found by us; one of them belonged to v. *gondaricum*, the other to v. *leucurum*. The remaining two chromosomes of the «quadrivalent» which are not connected with the nucleolus belong to some other pair.

The bivalent formed by the other pair of chromosomes with appendices in some cases was connected with the same nucleolus with which was connected the «quadrivalent»; however, there were cells with two nucleoli, and the bivalent was connected with the other nucleolus.

Inasmuch as the «quadrivalent» was formed only in those plants which possessed an asymmetrical chromosome set, i. e. whih had one small isobrachial chromosome with an appendix, one should think that it was just this chromosome of v. *gondaricum*, not the other larger one, which suffered reciprocal translocation and entered the «quadrivalent». Some well arranged figures prove this by making it clearly visible that out of the two chromosomes connected with the nucleolus one is larger and the other is smaller. Such a difference is observed between the remaining two chromosomes of the «quadrivalent», too.

Having established the structural reorganization of the chromosomes in one of the forms of the Abyssinian wheat, namely, in v. *gondaricum*, we shall pursue further the object to study the structural alterations in the other Abyssinian wheats as expressed in the reduction division. The data by Kostoff show that such alterations actually exist (2). This author has studied in one case the hybrids between *Tr. persicum* v. *stramineum* with *Tr. durum arraseita* from Abyssinia, in the other case—with *Tr. durum* from Mediterranean. In the latter case he has observed normal meiosis

whereas in the cross with *arraseita* polyvalents and univalents were observed. The author thinks that this fact may be interpreted along the line of structural difference between the two varieties of *Tr. durum*.

Most recently we have investigated rootlets of different Abyssinian varieties of *Tr. durum* which approach in their morphology and in their geographic distribution ([3]) v. *gondaricum*, viz., v. *axumicum*, *rubromuticum* and *taborense*. All of them had chromosomes with appendices similar to those of v. *gondaricum*, whereas var. *Schimperi* and *arraseita* differed in their chromosomes with appendices from v. *gondaricum*. This serves as an illustration of the nuclear polymorphism within the species *Tr. durum*.

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AGROPHYSIOLOGY

VARIATIONS IN THE VEGETATION PERIOD

By N. P. GLINYANY

(Communicated by A. A. Richter, Member of the Academy, 20. VII. 1939)

The investigations by Gregory and Purvis (1) as well as ours (2) have now established that at the time when the seed embryo is forming, the vernalization capacity of ripening seeds is in inverse proportion to their ripeness. The result is that phenotypically unhomogeneous progeny is produced from one and the same self-pollinating winter plant when after-ripening was carried out at low temperature.

The facts reported in the present communication are evidence that some unknown phenomenon is in co-operation with the factors responsible for the variable results obtained in spring-sowing of unvernalized winter plants.

Our experiments on wheat vernalization were carried on at Peterhof Biological Institute in 1937. Among the wheats investigated was *Triticum vulgare* Vill. var. *lutescens* Al., sort Harvest Queen. The seeds were supplied by the Central Storehouse of Gossortset. They were labeled as «Harvest 1936, U. S. A., Kansas». Part of them were vernalized and all were planted at two dates. The data characterizing the development of plants are summarized in Table 1.

Table 1

Variety	Vernalized during	Planted at	Start of heading	Expiration of heading	Number of days from planting till the outset of heading	Number of days from the start till the end of heading
Harvest Queen	46	1.V	1.VII	4.VII	61	4
	—	1.V	2.VII	—	62	—
	46	14.V	6.VII	8.VII	53	3
	—	14.V	10.VII	—	57	—

From these figures it is obvious that irrespective of whether the plants of Harvest Queen had received their start at the first or the second date, those of them which had been vernalized began heading only 1—4 days

earlier than those which had been not. Among the latter 8—15% came into ear during vegetation, the rest were tillering until winter set in.

The progeny from the unvernalized plants sown on May 14 1937 was obtained in 1938. Plants which had been the earliest to head were chosen for that purpose. Planting was done May 8. It was to be expected that now too the plants would come into ear without pre-sowing treatment given. Their parent plants had done so in the year 1937. Our expectation, however, was not supported by the facts and though the spring was cooler in 1938 than in 1937, the plants failed to pass natural vernalization and all of them without a single exception were tillering until late into the autumn. Still more surprising were the results obtained with part of the same seeds which after being made to germinate were planted in a cool greenhouse on April 14, and on April 29 were transferred in the open. Even plants of not too «wintery» kind should respond to such treatment by passing the full length of the vernalization stage. And so did actually the Spain winter wheat *albidum* which was planted at the same time (April 14), under the same conditions and successfully came into ear. But the progeny of the early plants of the Harvest Queen, despite the early time of planting, showed no sign of stalk formation until late in autumn.

The question naturally arises here, why should the unvernalized plants of the winter variety Harvest Queen which were able to head very early in 1937, almost simultaneously with the vernalized plants, produce in the following year progeny which remains in the stage of tillering throughout the vegetation period.

Lyssenko from his experimental data ([3]) arrived at the conclusion that by way of treatment with increased temperatures winter plants may become spring plants and remain such through following generations. According to a recently published paper by Avakian ([1]), the plants changed in that way are very thermophilous. Raised in a greenhouse they give a progeny which will not pass the vernalization stage in the field even if sown in spring. The explanation suggested by the author is that the low temperature prevailing in the field compared to that in the greenhouse does not meet the new requirements of hereditarily modified plants. In our experiments which were made on an abundant material uniform heading was recorded in the plants of Harvest Queen from the seeds vernalized during 46 days at 0—3°C, the results being similar for both dates of planting (Table 1). It will be obvious that reduced temperature failed to delay the development of some of the unvernalized plants sown in spring. These plants were able to come into ear alongside of the spring plants and they did so. It may therefore be said that the phenomenon observed in our experiments is different from what has been described by Lyssenko and Avakian.

We are lacking any ground to suppose that successful heading of unvernalized plants of Harvest Queen was due to their being vernalized during the period of embryo-formation, since it is hard to think that suitable conditions might be available in Kansas at the time of cereal ripening.

There were some similar facts more recorded in our experiments; let us describe them here too.

From 1935 until 1938 we were endeavouring to establish experimentally how the vegetation period in barleys is transmitted by heredity. It was necessary to reveal the diversity of forms in hybrid populations and we had recourse to the vernalization method. The results of those experiments will be described elsewhere. Here we shall deal only with some few facts bearing upon the individual development of the plants obtained from crossing a spring barley *Hordeum vulgare* var. *pallidum* Ser. sort Tchervonetz with a winter barley *Hordeum vulgare* var. *pallidum* Ser. (from Afghanistan).

A plant F_2, from which the progeny F_3-7301 was obtained in 1937, belonged to the set which had experienced a 43-day vernalization in the year 1936. It ripened very early and was harvested August 2. Among the progeny F_3-7301*28 plants were observed to tiller on July 11 and 33 plants were at the early shooting stage. Part of the plants was harvested September 19 while in the stage of «waxy» ripeness. Some of them produced F_4 families in 1938: the data on the development of plants in one of these families are referred to in Table 2. The figures show all the plants in the family to behave like true winter plants. But how to explain the fact that the parent plant F_3 produced ears in 1937 and was close to maturity («waxy» ripeness) on September 19? The seed it grew from could by no means have passed vernalization during the formation of its embryo in 1936, since it was ripe as early as August 2. Could it be that the plant (F_3) itself had suffered vernalization in the spring of 1937 after planting? But why had then the 96 plants F_4 failed to produce ears? Under conditions of a cool spring prevailing in 1938 they had no less frequently experienced reduction of temperature than had the parent plant in 1937.

Table 2

Planting done	Seedlings appeared above ground	Outset of heading	Expiration of heading	Data of counting the plants	Total number of plants	Produced ears	In the tillering phase	Data of harvesting F_3 in 1937
21.V	31.V	—	—	17.VII	96	—	96	19.IX
		—	—	28.VIII	96	—	96	
		—	—	1.IX	96	—	96	
		—	—	12.IX	96	11	85	

The meteorological data show that the seeds from which the family F_4 was grown were liable to vernalization during the period of ripening in 1937, but this cannot possibly be regarded as a factor of decisive influence.

The F_2 plant from which the Family F_3-7249 was obtained in 1937 was suffering vernalization in 1936 during 42 days. It was harvested on September 1. Family F_3-7249 began earing July 9. It included 144 plants of which 60 produced ears, 50 came into stalk, 24 were in the stage of early shooting and 2 in the stage of tillering on July 17. By August 27 112 plants of this family were harvested. From two of these families F_4-8823 and 8824 were obtained in 1938; they were tillering throughout vegetation. Family 8823 included 63 plants. 60 of them were tillering and 3 were at the stage of early shooting when inspected on September 24. Of 55 plants making up the family 8824 39 were tillering on September 8 while 16 were at the early shooting stage **.

No less surprising was the following phenomenon. A plant F_2 from a set of 1936 vernalized during 9 days was harvested September 21 at the early

* Planted May 16, 1937.

** It is of interest to note here that in certain families F_4 showing similar behaviour there were some few plants raised from seeds vernalized before sowing. These plants successfully came into ear while the unvernalized ones remained in the phase of tillering.

yellow-ripe and milk-ripe stage. From this plant 43 plants F_3 were obtained in 1937, 28 of which remained in the phase of tillering until 5. X. Part of the tillering plants was pricked off into boxes and kept in a cool room throughout winter. The aim was to harvest two crops during the season of 1938. On February 20 the plants were taken into a greenhouse varying in temperature from 12 to 22°C. During night they were supplied with electrical light from powerful bulbs. The plants reached the early shooting stage and in spite of continuous illumination began elongating very intensively. We therefore lowered the temperature to 4—12°C. The electrical light was continuous over every night. During March some plants sent stalks and some of them headed. The temperature in the greenhouse was then raised and kept between 17 and 24—27°C during April and May. Only occasionally did it fall down to 10°C. Consequently, the pollination and the development of the grain in our F_3 winter plants took place at high temperature and there was no possibility for the grain to be vernalized during the period of its embryo-formation.

The development of F_3 plants was delayed and we were no longer expecting the F_4 plants to yield in 1938. On June 25 there was planted about 60 unvernalized plants only. It was supposed that at least part of them would fail to head. Nevertheless, all plants F_4 produced ears and came into grain by August 26, though their parent plants had remained in the phase of tillering during nearly a whole year.

From thorough analysis of the facts here described we arrive at the assumption that the factors in the run of the vegetation period are not only the conditions which prevail when the plants are in the phase of forming embryos or later, but also those conditions under which the parent plants have developed. The reason for this is as follows. Vernalized plants yield progeny which has again to undergo vernalization. This points to the existence of some «critical process» in the individual life of plants which makes a vernalized plant set unvernalized seeds. According to our conception, it is by way of this «critical process» that the vernalization stage is establishing itself. Some external factors are certainly in operation here. If so, this process can be slowed down, strengthened or stopped through the use of external influences. But what are exactly these influences and at what interval in the development of the parent plant have they to come into play to make the critical process possible, is not yet known with any degree of certainty. All that can safely be said at the time being is that the critical process does not fall within the phase of vernalization; for it is just during this phase of plant development that the qualitative changes are brought about which are not transmitted to the progeny. For throwing light upon this intimate side of vegetable development it appears necessary above all to inquire into the rôle of the temperature in the formation of reproductive organs as well as pollen and egg-cells. Should the critical process take place during any of the phases within a given interval in the development of the parent plants, then, according to our assumption, the temperature conditions during this period might play a decisive part in determining the behaviour of the next (first) generation.

From analyzing the meteorological conditions which made the background for the development of plants in our experiments, we come to a conclusion which alone appears to be the most probable: the reason for the unexpected behaviour of plants described here should be sought in the temperature conditions, under which the development of parent plants has been taking place during the period from the end of vernalization until pollination.

It is not clear until now why should a non-homogeneous progeny be so frequently produced upon spring planting by the winter plants of one and the

same variety attaining maturity under condition of the south, where no vernalization at the stage of seed embryo-formation is possible during the summer months. Some plants among their progeny form ears and even ripen while others remain in the stage of tillering until the advent of winter. Our experiment with Harvest Queen proves that such a variation cannot always be explained in terms of genotypical unhomogeneousness. It is hard to believe that the differences in microconditions might be responsible for these great differences in the behaviour of plants. Perhaps, the responsible cause should be sought in the circumstance that in spring time (when temperature in the south is apt to vary within a wide range) different plants and even different stems of the same plant at various stages of this development (previous to pollination) are under the action of similar temperatures.

In our further experiments on the subject varying temperature conditions will be created for the vernalized winter plants to pass the various stages of their development (up to pollination). It will thus appear whether the critical process in question might be controlled (speeded up, slowed down or even stopped) by means of experimental intervention.

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PLANT PHYSIOLOGY

VERNALIZATION OF SMALL-SEEDED SUNFLOWER

By P. K. VOSHCHININ

(Communicated by N. I. Vavilov, Member of the Academy, 11. VII. 1939)

Duration of Various Phases in Sunflower's Life as Influenced by Vernalization, Photoperiodism and Environment. 1st Set of Tests. Vernalization at Low Temperature. Various phases of sunflower's life are controlled by vernalization, but all the individual plants do not respond equally to the treatment. In most of the cases the period from planting till the end of flowering becomes shorter while the subsequent phases of filling and after-ripening are prolonged.

1st, 2nd, and 3rd Experiments, 1934, 1935 and 1937. Original small-seeded variety Zelenka 22-82 acclimatized in Moscow since 1931 and vernalized by varying methods during 10, 20 and 30 days, at varying temperatures, on varying soils (heavy clayey ground of a high plateau in the 1st and 3rd experiment, low, swampy heavy argillaceous land with high water table in the 2nd experiment) and under varying climatic conditions of the three years above-mentioned invariably responded to vernalization at low temperature by shortening its life phases.

The phases under observation were budding, flowering and ripening from the beginning to the end. The decrease in duration for the various phases in 1934 ranged between 7.1 and 11.0 days; the period of ripening decreased by 7.9 days. The respective figures for the year 1935 are 2.4 to 6.9 days and 2.4 days; for the year 1937, 5.3 to 5.7 days and 5.9 days.

4th Experiment, 1936. An especially strong decrease in the duration of ripening—up to 18.2 days—was observed in a pure line of the Pioneer of Siberia. The seeds supplied by Voronezh Test Station were vernalized according to a simple procedure in a cellar in boxes at 5°C with pinching and pricking off the seedlings into an open border during the hot summer of 1936. The decrease in length of the other phases by this variety varied from 3.4 to 8.5 days.

2nd Set of Tests. Influence of Environment. a) The effect of vernalization varies according to the environmental conditions involving climate, microclimate, relief, soil, nutriment and supply of water. It may be mitigated or strengthened by these external influences.

5th Experiment, 1935. Experimental planting of variety Zelenka 22-82 was done May 26 in a high plateau clayey ground against control planting of the same population on May 19 in lowland podzol soil. Though vernalization was eliminated, the comparison showed a strong decrease in the

duration of ripening (10 days) and of other developmental phases (from 9.7 to 15.0 days) under the influence of external conditions only.

6th Experiment, 1936. Variety Saratov Early (second year of planting in Moscow) vernalized in a polythermostat at 4.9°C, 72% moisture, during 20 days and planted into various kinds of soils, under varying microclimatic conditions. May 20 into heavy argillaceous soil at Timiriazevka, May 23 into lean sand in the Moscow Zoo. The plants in Zoo ripened 12.2 days earlier, the other developmental periods being shortened by 2.4 to 7.7 days.

b) As a cold resistant and winterhardy plant the sunflower may be vernalized either by sowing it very late in autumn when the mean daily temperature is not above 3°C, or by sowing it very early in spring, in April, since it defies the spring frosts even at the phase of cotyledons.

7th, 8th and 9th Experiments, 1935. Variety Dwarf of Kherson acclimatized in Moscow since 1931, seeds from the harvest of 1934. There were three experimental plantings: two of them, with dry seeds planted very late in autumn (December 2, 1934) or very early in spring (April 21, 1935) and one, with seeds vernalized in boxes in the cellar at 6°C during 30 days. The control planting was done with soaked seeds May 19, 1935, in a swampy argillaceous lowland soil. The ultralate plants ripened 5.6 days earlier compared to controls and had their other life phases shortened by 5.4 to 6.0 days. The respective differences for the ultraearly plants are 5.2 and from 5.1 to 6.8 days; for the vernalized plants, 8.7 and from 7.9 to 9.7 days.

3rd Set of Tests. Data of Planting as a Factor Combined with Vernalization. Late sowing in itself is apt to reduce the developmental periods of the small-seeded sunflower, though the causes of this are not yet clear. The effect is still more pronounced, however, if late sowing is combined with low temperature vernalization. Sowing is late in the middle latitudes, if started with reference to the following phenological indices: when the dandelion has passed the height of its flowering season, after the horse-chestnut has finished flowering and the lilac has started it, after arrival of the martins.

10th and 11th Experiments, 1934. Populations of Zelenka 22-82 and Fuchsinka of Kherson acclimatized in Moscow since 1931; seeds from the harvest of 1933. Sowing of dry seeds on June 9 compared to control sowing on May 30 showed considerable reduction in all phases. Vernalization at 6°C strengthened the effect of late sowing (on June 9) and the various phases were still more reduced though by a variable number of days. Zelenka ripened 9.5 days earlier and finished flowering 12.8 days earlier, the difference in other periods varying from 7.7 to 8.2 days, when no vernalization was applied. The reduction added by vernalization was 0, 0.5 and from 3.6 to 4.5 days respectively. In the case of Fuchsinka the plants of the minus-vernalization variant ripened 3.8 days earlier, flowered 8.1 days earlier and the reduction of other periods ranged between 6.5 and 9.1 days. The effect of vernalization was further to reduce these figures: by 2, 0.8 and 0.9—2.9 days respectively.

12th Experiment, 1936. Variety Saratov Early. Two dates of sowing were compared against the background of a 20-day vernalization in boxes in the cellar at 6 and 8°C with pricking off into boxes and then in the open ground. The boxes were moved into light on May 30 and June 9; the phase of ripening only was delayed in the late-started plants because rainy weather set in. In the early variant the plants grew ripe while drought prevailed and there was only one rain on June 16 which gave 40 mm of precipitation. The decrease in duration was 0.2 for ripening; 6.0 for the end of flowering and from 3.3 to 6.1 for the other phases.

Influence of Pinching Combined with Vernalization. Similar phenomenon is observed when sunflower's seedlings are pinched into boxes with humus and pricked off in the ground after they had taken root. Here, too, the decrease in the duration of various phases receives added strength from vernalization.

13th Experiment, 1937. Population of Zelenka 22-82 from seeds harvested in 1936. Vernalized in boxes at 7.3° C, 80% moisture, during 20 days. The test for pinching was run with strongly sprouted seeds planted into boxes. Pinching when applied singly reduced the ripening period by 4.9 days and the other phases by 3.4 to 4.4 days. When it was combined with vernalization further decrease by 0.6 and 1.7 to 1.9 respectively was observed.

Vernalized Plants under the Influence of Photoperiodism. The result of a 10-hour day photoperiodic treatment of plants vernalized into boxes was to increase the effect of vernalization by decreasing the length of life phases farther still. Varying degrees of reduction were obtained in the experiments with various varieties of sunflower. One single exception was the variety Saratov Early. The duration of its developmental phases was found neither to decrease nor to increase in response to photoperiods and only showed insignificant fluctuations in both directions.

14th Experiment, 1937. Variety Pioneer of Siberia. Seed harvested in 1936 were used in planting. Vernalization in boxes in a cellar at 7.3° C during 20 days was supplemented by photoperiodic treatment with 10-hour days, and the cumulative effect was strongly to reduce the life periods. Ripening was shortened by 4.4 days, the decrease in the length of the other phases varied between 1.1 and 5.0 days.

Other Modifications in Small-Seeded Sunflower Induced by Vernalization, Photoperiodism and Pinching. 4th Set of Tests. The response of the small-seeded sunflower to the vernalization, photoperiodic treatment and pinching combined with the influence of the environment at varying temperatures is to shorten the length of the stem and to decrease or increase the diameter of the head and the yield of seed. Vernalization at high temperature protracts the life periods.

Among the combination of external influences which in their action are linked up with vernalization, the influence of nutrient substances of the soil is especially conspicuous, though it is not directed to shorten the developmental phases.

15th Experiment, 1935. Population of the variety Dwarf of Kherson. Lean soil of a heavy podzol lowland. Seeds vernalized at 6° C fell upon a spot of more fertile soil: all developmental phases increased in length; the height of the stem went up from 72.0 to 83.3 cm. It will be obvious that here the effect of vernalization was offset, the life periods increasing in duration by 0.4 to 2.1 days; but then was also observed a strong increase in the head diameter, from 11.0 to 12.2, and the yield of seed from the middle head was almost doubled (increased by 96.4%). In the three other experimental series on lean soils the effect of vernalization was as usual: the height of the stem and all developmental phases were reduced in length, but the size of the heads and the yield of seed were also lower.

16th Experiment, 1936. Saratov Early was planted under condition of drought into lean sand among sun-heated stone buildings of the Moscow Zoo. A new combination of modifications due to the action of vernalization was here obtained. The flowering phase grew 1.7 day longer while the other phases were slightly shortened; along with this the length of the stem decreased by 0.3 cm, while the diameter of the head became longer by 0.5 cm,

and a strong increase of the yield (by 31.4%) took place in spite of obvious deficiency in nutriment and moisture.

17th Experiment, 1936. The same population of variety Saratov Early was grown on heavy argillaceous soil at Timiriazevka under more favourable environmental conditions, after having been planted late and enjoyed optical temperature during vernalization (6—10° C). Its response to vernalization was positive according to all indices, and the yield increased by 22.5%. The duration of the ripening period decreased by 2.7 days; the reduction of the other phases varied between 0.5 and 2.9 days; the stem was shorter by 2.3 cm, while the diameter of the head was 0.8 cm longer.

Attention should also be given to the extreme sensitivity of the sunflower to short and abrupt changes in temperature during vernalization. This phenomenon is worth special investigation.

18th Experiment, 1936. A population of the variety Fuchsinka of Kherson enjoyed optimal temperature during vernalization (4.9° C) and responded positively to the treatment according to all indices, increasing moreover its yield by 6.6%. If however the optimum temperature of vernalization was interrupted for 2—3 hours by increasing it sharply to 25—30° C and this repeated every 5 days, 4 times within 20 days, the reduction of all indices was the result and the increase in yield became 5.4% lower.

The effect of short (10-hour day) photoperiods was always to reduce all indices and the yield of seed in all of the populations treated. The only exception were the selected seeds of variety Saratov Early. This variety did not reduce its life periods in response to short days.

The stem was invariably observed to decrease in length normally under varying conditions of vernalization, photoperiodism and pinching.

The diameter of the head grows longer in response to vernalization carried on at a low or high temperature under optimal external conditions.

If the sunflower is vernalized at a high temperature, a particular reaction is observed: the life phases are reduced, but the stem grows shorter, the head broader and the yield of seed is increased.

19th Experiment, 1937. A population of the variety Pioneer of Siberia, vernalized at 25°C during 20 days in salt solutions to inhibit growth, was observed to lengthen its ripening period by 13 days.

The mean height of the stem was reduced from 119.0 to 114.2 cm. The head increased in diameter by 0.1 cm. The yield was 99.8% higher compared to control. Similar results were obtained with variety Zelenka 22-82: reduction of stem height from 124.6 down to 115.6 cm; increase in head diameter, 1.0 cm.

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ANIMAL PHYSIOLOGY

ENERGY BALANCE OF THE GROWING LARVA OF *SILURUS GLANIS*

By V. S. IVLEV

(Communicated by I. I. Schmalhausen, Member of the Academy, 27. VII. 1939)

The study of energy exchange in a growing animal has led to a number of important generalizations [(7,9) and others]. Data are available indicating the nature of energy processes during embryonic growth (2, 6,8). We have made attempts at tracing age variations in some of the energy indices of invertebrates (4) and determining the basic elements of the energy balance in growing fish (5). Of considerable interest is the question of energy transformation in fish larvae though living under field conditions, but existing and growing at the expense of internal energy reserves (yolk).

The work has been performed on *Silurus glanis* larvae which owing to the size, quality of the yolk-sac and endurance present excellent experimental material. The experiments have been carried on from the moment of hatching from the egg until final resorption of the yolk. The growth of larvae and the degree of yolk absorption were determined from the calorimetric analysis of larvae made every 24 hours. The yolk-sac was removed in larvae fixed with 10% formalin. Caloric value was determined by the method of wet combustion (3). From 40 to 60 specimens were burnt for each age. Gaseous exchange was determined by means of a Drastich (1) microrespirometer. Excretion energy was calculated on the basis of amounts of nitrogen excreted by 100 larvae within 24 hours and determined by the microkjeldhal method.

Table 1

Age (in days)	Number of specimens tested	Amount of O_2 in mm^3 (per 1 mg. of dry weight per hour)	Amount of O_2 in mm^3 (per specimen per hour)	Caloric value of respiration in g-cal. (per specimen per hour)
0—1	250	6.32	0.743	0.00334
1—2	250	6.24	0.814	0.00366
2—3	250	5.43	0.842	0.00379
3—4	150	5.51	1.350	0.00607
4—5	150	5.36	2.198	0.00989
5—6	150	5.22	3.341	0.01503
6—7	150	5.07	4.690	0.02110
7—8	150	5.09	5.777	0.02600

Table 1 shows the results of energy measurements of external and internal work which are inseparable as far as our material is concerned, the larvae of *Silurus glanis* being in a state of perpetual oscillatory motion. To express the volume units of oxygen in calories, on the basis of the chemical composition of yolk, we are using the oxicaloric coefficient equal to 0.0045. All data on the energy balance are summarized in Table 2.

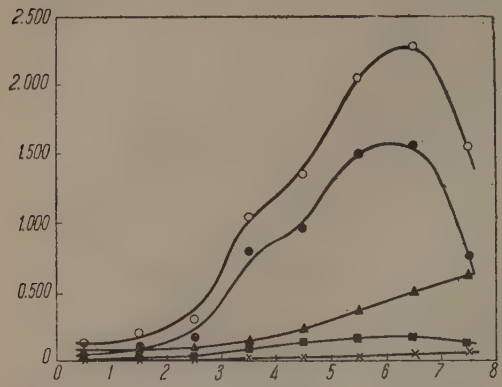

Fig. 1. Energy balance of one larva. Abscissa, age in days; ordinate, gram-calories. Light circles—Q (energy consumed), shaded circles—Q', triangles—Q_{w+v}, squares—Q_t, crosses—Q_R.

Table 3 (Fig. 2), where the total amount of the energy expended is taken to be 100%, presents a better illustration of the dynamics of the energy balance.

On the basis of the values derived have been calculated the coefficients of energy consumption computed by us (l. c.) by means of the fundamental equation of Terroine and Wurmser (9).

Table 4 shows the dynamics of these coefficients during the process of growth of larvae and mean coefficients for the whole larval period.

Table 2

Age (in days)	Caloric value of yolk (in g-cal.)	Caloric value of yolk absorbed per day (in g-cal.)	Caloric value of one larva (in g-cal.)	Caloric value of increment of one larva per day (in g-cal.)	Caloric value of excretions by one larva per day (in g-cal.)	Caloric value of respiration per day (in g-cal.)	Amount of initial heat (in g-cal.)
0—1	8.7483	0.1353	0.6328	0.0389	0.0066	0.0802	0.0096
1—2	8.5402	0.2081	0.7336	0.1008	0.0073	0.0878	0.0122
2—3	8.2479	0.2923	0.896	0.1624	0.0078	0.0910	0.0311
3—4	7.2051	1.0428	1.686	0.7900	0.0122	0.1457	0.0949
4—5	5.8631	1.3420	2.640	0.9540	0.0204	0.2374	0.1302
5—6	3.8159	2.0472	4.134	1.4940	0.0311	0.3607	0.1614
6—7	1.5393	2.2766	5.692	1.5580	0.0483	0.5064	0.1639
7—8	0	1.5393	6.432	0.7400	0.0517	0.6240	0.1236
0—8	—	8.8836	—	5.8381	0.1854	2.1332	0.7269

Table 3

Age (in days)	Caloric value of increment in % (Q')	Caloric value of excretions in % (Q_R)	Caloric value of respiration in % (Q_{w+v})	Initial heat in % (Q_t)
0—1	28.75	4.87	59.28	7.10
1—2	48.44	3.51	42.19	5.86
2—3	55.56	2.67	31.13	10.64
3—4	75.76	1.17	13.97	9.10
4—5	71.09	1.52	17.69	9.70
5—6	72.98	1.52	17.62	7.88
6—7	68.44	2.12	22.24	7.20
7—8	48.07	3.36	40.54	8.03
0—8	65.71	2.09	24.01	8.18

Analysis of the materials presented in Tables 3 and 4 suggests the following characteristic features of the growth dynamics of fish larvae.

1. The coefficients of energy consumption during the growth of fish larvae are considerably higher than those in all actively fed animals.

2. Comparatively low coefficients of the first and second order during the first and partly second and third 24-hour periods are due to the extremely

Table 4

Age (in days)	Coefficient of energy consumption of first order	Coefficient of energy consumption of second order	Coefficient of energy consumption of third order
0—1	28.75	30.23	80.21
1—2	48.44	50.20	89.20
2—3	55.56	57.08	83.93
3—4	75.76	76.65	89.28
4—5	71.09	72.19	88.07
5—6	72.98	74.10	90.25
6—7	68.44	69.92	90.43
7—8	48.07	49.74	85.69
0—8	65.71	67.12	88.78

high level of the basic metabolism (respiraton energy). The same holds good for the terminal larval stage.

3. The coefficient of energy consumed of the third order is subject to irregular changes within rather narrow range (from 80 to 90%).

4. The energy of waste-products is very deficient testifying to the organism being extremely economical in spending the internal energy reserves.

5. The factor limiting the growth rate of the larva is not the degree of yolk absorption ([2]), but the intensity of metabolism.

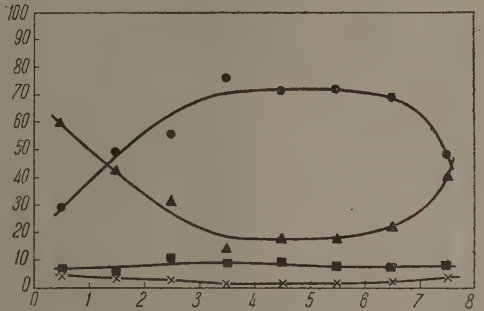

Fig. 2. Variations of energy balance with age. Abscissa, age in days; ordinate, percentage relation of energy expenditures. Symbols —same as in Fig. 1.

6. The dynamics of growth of *Silurus glanis* larva is, according to the nature of food, a typical instance of energy transformation during the embryonic development proper.

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ANIMAL PHYSIOLOGY

EFFECT OF STARVATION ON ENERGY TRANSFORMATION DURING THE GROWTH OF FISH

By V. S. IVLEV

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The degree of energy expenditure during the growth of an animal is quite a constant value, but little affected by external or internal factors, subject to variation when transferred to a mainly qualitatively different ration and decreasing with growth of the organism (2). Considerable interest therefore belongs to the data by Matsui and Oshima (5), who report of a carp having showed after starvation a higher increment per unit of foodstuff than the one regularly fed.

In the present work the author attempts to analyze this phenomenon and disclose the phases of energy metabolism which are the first to be affected by previous starvation. The work has been carried out on the young of pike *Esox lucius* which is the most suitable object for experimenting with growth of fish (6,4).

The following procedure has been applied. The experiments were performed at a temperature of 16—18°. The young hatched from artificially fertilized eggs were normally fed, whereupon 10 specimens were taken for regular investigation of energy metabolism, while the rest were left without food. After 10 days of starvation another 10 fish were selected, then followed others after 20 and 30 day intervals. On the 34—35th day the experimental fish began to die in mass.

After starvation the fish were given abundant food (young of *Rutilus rutilus*). The amount of food consumed was figured from difference between that previously weighed and the remainder. A parallel analysis of the food made it possible to express weight units in calories. The excretion energy was calculated from the amount of nitrogen excreted in the form of products of protein metabolism determined by the microkjeldahl method and amount of excrements combusted in chromium compound subsequently expressed in calories (1). The respiratory level was determined after Krogh. The experiments were continued until the fish of the given series have doubled their weight. Analysis of the material under experiment before and after the observation showed the increment for the period concerned.

Table 1 shows figures (means from 10 parallel experiments) characterizing the results of starvation and indicating the time required by the fish for doubling the weight after starvation.

Results of determination of elements of the energy metabolism are recorded in Table 2. In this case each figure is the mean obtained for 10 experi-

Table 1

Starvation period (in days)	Initial size of young (in g-cal.)	Size of young after starvation (in g-cal.)	Loss during starvation (in g-cal.)	Loss during starvation (in %)	Time required for doubling the weight (in days)
0	1195	—	—	—	12
10	1207	996	211	17.5	10
20	1203	894	309	25.7	9
30	1189	822	367	30.9	14

mental objects. Determinations of the respiratory level and excretion energy were made every 48 hours and then integrated for the whole period of observation. The quantity of initial heat was calculated from the difference between the amount of energy expended and the other expenditures.

Table 2

Starvation period (in days)	Absorption energy Q (in g-cal.)	Increment energy Q' (in g-cal.)	Excretion energy Q_k (in g-cal.)	Energy of external and internal work Q_{w+v} (in g-cal.)	Quantity of initial heat Q_t (in g-cal.)
0	3780	1190	1018	377	1195
10	2806	986	698	271	851
20	2348	906	525	212	705
30	3089	802	1063	243	981

From the values given in Table 2 are calculated coefficients of energy consumption for all of the 4 series (Table 3).

Table 3

Starvation period (in days)	Coefficient of energy consumption of the first order	Coefficient of energy consumption of the second order	Coefficient of energy consumption of the third order
0	31.48	43.08	49.90
10	35.14	46.77	53.67
20	38.59	49.70	56.24
30	25.96	39.59	44.98

It may be seen that the energy indices of a normally fed pike are nearly equivalent to the figures obtained for a carp dwelling in natural waters ([3]). After a certain period of starvation the energy consumed is more adequately utilized, whereby a rise of coefficients of energy expenditures of all orders is taking place, pointing to the ability of the organism to save energy throughout all the phases of the energy balance. Let us note that the increase of coefficients stands in definite, almost direct dependence not upon the amount of substance lost during starvation (Table 1), but upon the length

of starvation. This condition is presented in the diagram where along the axis of abscissae is plotted the starvation period (in days), and against the axis of ordinates—coefficients of energy consumption.

The above is true except in cases of prolonged starvation. As shown above, mass destruction began on the 34—35th day of starvation. Beginning from the 32 day autolysis of proteins set in. Having starved for 30 days fish bore all the marks of profound starvation: nearly complete cessation of movement, rare respiratory rhythm, deeply sunken eyeballs etc. Starvation among fish of this series reached such a degree that upon resumption of feeding the fish but regained slowly the normal metabolism, their basic energy functions having been upset for a long time.

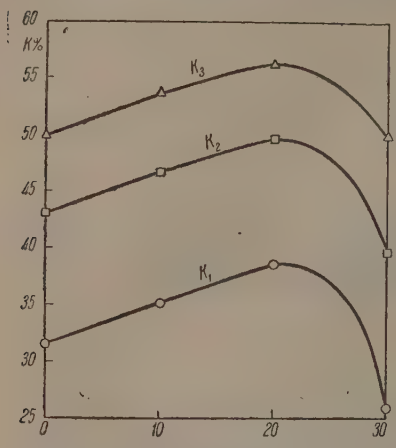

Table 4 illustrates the changes in the energy balance caused by previous starvation, the elements of energy exchange being expressed in percentage of the amounts of energy expended for the same period.

It follows from this table that starvation of the young of pike not exceeding 20 days causes the energy balance to shift toward a higher rate of expenditure of the energy consumed for the process of synthesis of plastic substances (resp. growth) and more rational use of energy for all other purposes.

Table 4

Starvation period (in days)	Increment energy Q' (in %)	Excretion energy Q_R (in %)	Energy of external and internal work Q_{w+v} (in %)	Amount of initial heat Q_t (in %)
0	31.48	26.93	9.98	31.61
10	35.14	24.86	9.61	30.39
20	38.59	22.35	9.02	30.04
30	25.93	34.42	7.88	31.77

It is noteworthy that even the rate of energy output in the form of initial heat, being the result of a definite succession of reaction of disintegration and synthesis, decreases to a certain extent.

Profound starvation, on the contrary, results in decreased percentage of energy expended on the process of growth. The principal cause of this phenomenon is a serious upsetting of the process of digestion, accompanied by increased elimination of the undigested food particles.

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PARASITOLOGY

NUTRITION OF TAPEWORMS IN ARTIFICIAL MEDIA

By G. S. MARKOV

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A study of helminths' nutrition in vitro is indispensable for their successful culftivation in rationally composed artificial media and will help the understanding of parasites' nutritive biology in vivo. The effect produced by helminths in the host animal can be often due to the method by which parasites obtain their food.

Tapeworms (*Cestodes*) are the most favourable objects for such a study, because they «... lack a mouth and digestive tract and must absorb all of their food through the surface of their body» ([10]). Most of the investigators ([5, 6, 10, 11, 16-18]) suppose that tapeworms live predominantly upon glucose. The literature ([1,4,17]) on the nutrition of tepeworms (*Moniezia*, *Diphyllobothrium*) in artificial media is very scanty. The defect of these works is the short duration of the experiments (6 hours). The author ([9]) succeeded to keep some fish helminths alive outside the body of their hosts in artificial conditions for 30—60 days. This opened the possibility to investigate the nutrition of parasitic worms in artificial media. The procedure and some results of the cultivaton have already been described ([9]).

There is a general agreement among parasitologists ([2, 7, 10, 18]) that glycogen is a fundamental energetic source of the anoxidative metabolism of parasitic worms. According to this interpretation, the glycogen accumulation in worms and its dynamic in various media were chosen as a criterion to evaluate the starvation or nutrition of the parasites in these media. The amount of glycogen was determined chemically by the method of Pflüger-Osterberg. The reducing sugar was estimated by the method of Iwataka with modifications of Kochneva and Kazimirova. All analyses were made twice. The influence of the variation in glycogen content among individual tapeworms was eliminated owing to the abundance of experimental material. Glycogen was also studied morphologically under the microscope (in conventionally adopted categories) on thousands of sections stained by the method of Best. The data were biometrically analyzed. Both methods gave similar and comparable results. The results of chemical determinations of glycogen content are given in the table.

The initial glycogen content in the helminths' body is very high; they contain 4—10 times more glycogen, than their hosts. For example, the glycogen content of plerocercoids of *Ligula intestinalis* is on the average 33.7%, that of *Eubothrium rugosum*—29.8%, of plerocercoid of *Diphyllobothrium latum*—17.9%, and the glycogen content of *Triaenophorus nodulosus*

is 13.8% of the dry weight. The glycogen content of muscles of the pike (*Esox lucius*), their host, is 3.4% and that of the liver only 27.1% of the dry weight. Such relations were already found in mammalia helminths by Smorodinzev and Bebeshin ([12],[13],[14],[15]) and others. Glycogen granules are distributed uniformly throughout the parenchyma and eggs.

Glycogen Content of Parasitic Worms in Various Media According to Data of Chemical Analyses

Species	Media	Duration of experiments in hours	Mean glycogen content in g % of dry weight	Rest of glycogen in % to initial amounts
1. Plerocercoids of *Ligula intestinalis*	Initial	—	33.7	100
Do	Ringer-Locke	336	23.1	70
»	Ringer-Locke+glucose 1%	280	34.7	103
»	Ringer-Locke+fish broth	528	37.4	111
2. *Eubothrium rugosum*	Initial	—	22.8	100
» »	Ringer-Locke	192	22.5	76
» »	Do	720	10.0	34
» »	Ringer-Locke+glucose	216	30.3	100.2
3. Plerocercoids of *Diphyllobothrium latum*	Initial	—	17.9	100
Do	Ringer-Locke; 15°C	38	11.9	66
»	Ringer-Locke; 35°C	38	3.9	22
»	Do	72	1.0	6
»	Ringer-Locke+glucose, 35 C	38	18.5	103
4. *Triaenophorus nodulosus*	Initial	—	13.8	100
» »	Ringer-Locke	264	8.8	60
» »	Ringer-Locke+glucose	264	15.3	111

In Fig. 1 are represented results of the morphological examination of glycogen during 720 hours in plerocercoids of *D. latum*, in Fig. 2—data of chemical determinations in the same object. In Ringer-Locke plerocercoids have no food, their mobility there is the highest. Accordingly, they starve and consume their endogen-storage of the energetic material—the glycogen. Fresh plerocercoids show an amount of glycogen corresponding to the category 3.6; after 720 hours this amount decreases to the category 1.6. The diminution is statistically significant after 180 hours of starvation. In glucose and in fish broth the glycogen content of worms diminished slower (see both curves in Fig. 1). During 480 hours the plerocercoids' glycogen content in these media corresponds to the category 3, i. e. is by a whole category higher than in «starving» worms.

The same conclusions might be drawn from the data represented in Fig. 2. After immersion of plerocercoids in Ringer-Locke for 38 hours, they consume 34% of their glycogen. If the glycogen content of «Ringer's» worms after 120 hours makes up only 65% of the initial one, that of «glucose's» plerocercoids is 85% and of «broth's» ones 95%. After 240 hours—57, 76 and 82%, respectively. When after 360 hours starvation in Ringer-Locke the rest of glycogen is 17%, plerocercoids from glucose have three time more (57%) and those from the fish broth five time (79%) more glycogen. The

degeneration phenomena after 480 hours cause a sharp decrease in the glycogen content (Fig. 1 and 2). It should be emphasized that fish broth contains food (perhaps glucose) enough for plerocercoids. Ringer-Locke glucose solution is less favourable in the physical-chemical sense than fish broth, and plerocercoids lose there more glycogen.

These data are fully confirmed by experiments with others worms (table). The plerocercoids of *Ligula*, when kept in Ringer-Locke for 336 hours consume 30% of the initial amount of glycogen; when glucose or fish broth are added, no diminution takes place. The difference observed depends on the nutrition of worms in these nutritive media, because the rate of movement of worms was the same. The rest of the glycogen content in *Eubothrium rugosum* after 192 hours immersion in Ringer-Locke is 76%, but when glucose or broth are added no consumption of glycogen could be observed. In *Triaenophorus nodulosus* the glycogen content diminishes (60% of the initial), when worms starve in

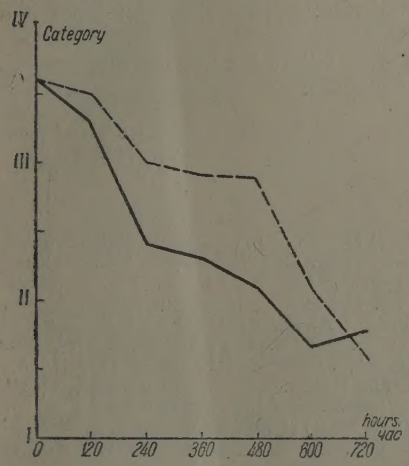

Fig. 1. The dynamic of glycogen of plerocercoids of *D. latum* in Ringer-Locke (——), glucose (— — —) and fish broth (—·—·—) (morphological data).

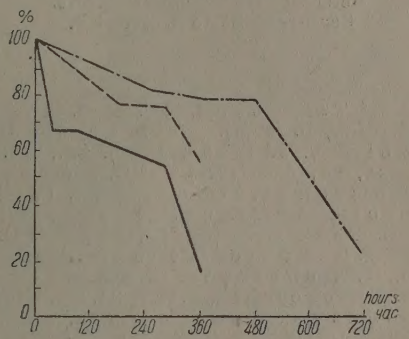

Fig. 2. The dynamic of glycogen amounts (in % to the initial) of plerocercoids of *D. latum* in Ringer-Locke (——), glucose (———) and fish broth (—·—·—) (results of chemical determinations).

Ringer-Locke; «glucose» tapeworms show no variation in the glycogen content. Histological control in all cases confirms these conclusions.

If we compare the level of initial amounts of glycogen in various species with the rate of its consumption, it may be seen that the rate does not depend upon the initial level as it is the case in mammals. For example, the glycogen content of plerocercoids of a broad tapeworm is 17.9% of the dry weight, the consumption during 360 hours being 83% of the initial amount. *Ligula's* plerocercoids show a glycogen content of 33.7%, consuming during 336 hours only 30% of the initial amount. In ascarids this fact had been reported by Brand(3).

The rate of glycogen consumption is dependent upon the mobility of worms and their metabolic activity. The rate of movement of the anterior region of plerocercoids of *D. latum* is at 15° C 5.5 expansions and contractions per minute. Glycogen consumption for 38 hours—34% of the initial (table). At 35° the rhythm is 16.9, the glycogen consumption—78%. After 72 hours of starvation at 35° the glycogen rest is only 6% of the initial. When glucose is added, no diminution in glycogen could be observed.

Significant results were also obtained from quantitative determinations (method of Bertrand) of the glucose consumption in media with worms. For

example, plerocercoids of *D. latum*, being immersed in 0.4% glucose solution, consume for 100 mg per 24 hours 0.493 mg of glucose; in another experiment with 2% glucose—0.812 mg. Plerocercoids of *Ligula* under the same conditions consumed 1.909 mg of glucose. The low level of the glucose consumption in comparison, for example, with that of tissue cultures ([8]) is obviously due to the high glycogen content in helminths at the first days of the experiment.

The essential results of the work are:

1. Experiments with starvation confirm that glycogen is a fundamental energetic source of parasitic worms.

2. Rate of the glycogen consumption of starving worms is independent of the initial glycogen content and depends upon the intensity of the metabolism and the rate of movements of helminths. Both factors are dependent upon the temperature of the media.

3. My experiments show beyond any doubt that tapeworms may absorb as food nutritive materials of the kind of glucose.

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